INFORMATION COMPUTERS MACHINES AND HUMANS

Edited by:

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Information, Computers, Machines and Humans

An Introductory Text to Systems Engineering

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An Introductory Text to Systems Engineering

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PREFACE

This book has been written largely in response to requests from students who, while preparing for a course at the University, would like to have an overview of methods, powers and limitations peculiar to modern engineering and technology, with some attempt at an estimate of the effects which such developments might have on human societies.

The material for this book grew out of notes prepared for lectures given to first year students: the whole course consisting of about 30 lectures. In the space provided it would, therefore, be impossible to present a substantial proportion of the material needed for a technical discussion of systems engineering and the technique of modelling.

Instead, various facets of systems engineering are unfolded in stages, but above all, an attempt is made to help the reader to appreciate the methodology and the way of thinking peculiar to systems engineering. The modern engineer with the vast technological and financial resources at his disposal, can no longer consider a machine in isolation, but rather he must view the machine in its environment, and include in his estimate, the likely effects which the machine might have on human beings.

The book is divided into 20 chapters dealing with (1) Information, (2) Engineering materials, (3) Building blocks and anatomy of systems, (4) Computers, (5) Systems (such as communications), (6) The human element, and finally a reference to (7) Global systems.

Discussion on information is fundamental to understanding of the system's function: information in various disguises of digital data or electric waveforms is the life-blood of every system without which it could not perform its functions. The information network of a system can be likened to the nervous system of a living organism; both as regards the functions performed, as well as its complexity.

A system is made up of a variety of building blocks and functional units in accordance with a pre-arranged plan, but ultimately the performance of a building block depends on the fundamental properties of the engineering materials used. Indeed, it would be true to say that the progress with large systems which we have witnessed in the last decade, would not have been possible but for the incredible developments which have taken place in the technology that enables us to produce ultra-pure materials and to control their properties to an extraordinary degree of accuracy. This progress led ultimately to the mass-production of reliable micro-miniature and integrated electronic circuits accompanied by considerable reduction in cost. In fact, in many cases, we have reached the ultimate limit set by the molecular structure of the materials.

The progress in materials science has been paralleled by an equally spectacular progress in systems (this includes theory, methods and hardware accomplishments). Currently systems are planned to such a high degree of complexity that no human being can comprehend its detailed *modus operandi* or be capable of using it in an optimum fashion.

For example, with large electronic computers, we can put to an effective use only a fraction of their potential power. With many a large system, such as a space project, the costs of a mission accomplished often turn out to be orders of magnitude higher than the original estimates. (It is not uncommon to find that a project which has been estimated at around $10m reaches total costs in excess of $1000m.)

The evidence to date is abundantly clear: we do not understand how to design, plan nor manage efficiently a large system. The appalling fact remains that many a multi-million dollar decision is made substantially on an *ad hoc* basis with scanty evidence to support the actions and in the face of a complete lack of understanding of any possible repercussions on other systems or human societies (or even the very existence of the human race). It is for these reasons that the chapter on human systems has been included in this volume.

How shall we prepare ourselves for the even larger and more numerous projects of tomorrow? We need foresight and courage: foresight to see the problems ahead and courage to tackle them. Greater emphasis must be put on co-operative efforts. (Indeed this book itself is the result of a co-operative effort). Hence the need for more and better education.

The final chapter is devoted to some philosophical issues relating to systems engineering. It attempts to summarize some of the principal issues and terminates on a note of hope for humanity: the key to the solution of the many problems of tomorrow is really in our hands. To a very large extent it is *we* who determine the shape of tomorrow, the ugliness is unlikely to be *their* fault.

The technological problems evolve around patterns and semantics: how to translate semantics into patterns amenable to scientific scrutiny is really the issue behind many a major technological problem.

Systems engineering is really an attitude, a way of thinking, a method of putting a major plan into operation. But, in the relatively narrow sense of present day concepts, it is likely, in the long run, to create more problems than it will solve. We can already see the need for a change in attitude.

Thanks to the developments in systems engineering, we can now deal with projects of magnitude hitherto undreamed of, so large in fact that should the project result in a failure, a whole nation could be ruined. Furthermore, the operation of a large project could modify the environment fundamentally, and unless this too can be taken into account, the system could lead to disaster.

The only satisfactory solution seems to be in the introduction of a "global system" in which we have the fundamental obligation to ensure that any new system to be introduced, as well as the progress and the functioning of monitoring the operation of others already existing, is under the control of a global system performing in the light of a genuine concern for the well being of every nation and every individual.

To be educated means to have a plan for tomorrow with a genuine concern for others.

A.E.K. & R.M.H.

"Living is an art and, to practice it well, men need not only acquire skill, but also tact and taste".

(Aldous Huxley).

CHAPTER 1

On Patterns, Systems and Creation

"What is this life if, full of care,
We have no time to stand and stare.
. . . . .
No time to turn at Beauty's glance,
And watch her feet, how they can dance."

W. H. Davies.

CHAPTER 1

On Patterns, Systems and Creation.

1.1 Introduction

Applied Science – and Engineering in particular, represents, perhaps, the largest and the most rapidly growing volume of knowledge and human skills. At the same time, our progress is almost entirely dependent on developments in technology.

It is now recognised that well developed technology is not only necessary for the well-being of every country, but that technology is the key to success, certainly a key to power, or a key to destruction if we choose so. But technology is also a creative way of life, a way to self-fulfilment, to achievement, to greatness, as well as to prosperity.

Engineering is also a discipline in its own right, with its own tradition and folklore. But whereas technological knowledge and skills grow in complexity and diversity, the work of an engineer is also essentially creative with intense artistic overtones, and it is these overtones with their counterpart of abstract patterns in our minds, that are a source of immense pleasure and impart a sense of professional achievement.

Engineering has been undergoing great changes in the last few decades. These stem from two directions:

(i) a growing profusion of specialised disciplines, and

(ii) the emergence of very large projects requiring large scale team work of specialists, with associated human problems.

Thus a snow–ball effect with all its problems has been established: A large project, for its solution, requires greater specialisation in narrower fields, leading to the need of a larger number of specialists and greater difficulties with team work, thus accentuating the magnitude of problems associated with large projects. Let us now examine briefly how this has come about.

Firstly, we observe that over the years we have come to differentiate between various branches of knowledge and activities involving applied sciences and engineering. Thus, very specialised disciplines have evolved under different headings such as oceanography, communication engineering, management, medical electronics, computers, satellites and many, many others. This profusion of branches is natural to the development of our knowledge. As the years progress, we learn more and more about the various branches of knowledge, and the various compartments grow in size accordingly. As a result, in most professions we tend to specialize, and specialize we must if progress is to continue.

More recently, particularly since the second World War, there has been another development taking place simultaneously. Here we have in mind the emergence of very large engineering projects, such as satellite communication systems, particle accelerators for scientific research, complex traffic systems, large power distribution networks, completely automated factories and many other examples. What is peculiar about such projects is *not* new, neither involved nor unusual scientific phenomena, but rather the complexity which arises from the sheer size of the project. A large project brings new problems: various forces which had negligible influence on the performance of a particular component can play important roles when such components become (in great numbers) parts of a complete system. Large systems are invariably expensive and the safety factors which one tends to impose are more stringent than those for smaller projects. Finally, large projects contain many inter-connected components and sub-assemblies; it thus becomes difficult for a human being to comprehend the functioning of the system as a whole.

These two factors, extreme degree of specialisation and the emergence of large systems of great complexity (which are difficult to comprehend), produce far reaching human problems. Firstly there is the problem that no single person can comprehend adequately all matters connected with a large system; no one can, therefore, be said to be wholly in control of it. Team work at all levels, including the uppermost level of management becomes mandatory. This implies a change in attitudes of specialists, managers, etc. from a self-centered to a task oriented motivation.

We thus come to the curious conclusion that further progress – which includes fundamental scientific discovery and which was in the past largely a function of self-centered and self-motivated inventiveness (backroom boy approach) – is now recognised to be limited by the rate of progress of technology which we can maintain, and this in turn is primarily limited by human problems. It, therefore, becomes imperative to gain a better understanding not only of the complex system in isolation, but the system in which machines and humans interact on a vast scale in a common environment. How then can we deal with the design and management of large systems?

First of all we must try to answer the questions. What is a "System"? What is "System Engineering"?

Clearly a system* is man made. It is large and complex, made up from numerous parts which perform various functions. The components of a system are inter-connected to form functional units and these in turn are further connected to form a complete assembly, with suitable input and output units. The various parts of the system may be independent, but are capable of being separated and inter-connected in a variety of patterns, but above all the whole assembly is designed to function according to a definite plan and performs as an integral unit.

Usually it is not possible to "design" a system in the ordinary sense of the word. One, therefore, builds various mathematical models and prototype working models before the design stage can be reached and even then the system is equipped with self-adaptive features to enable it to make decisions in new circumstances.

The self-adaptive and decision making characteristics make the system's behaviour difficult to predict, but at the same time make the system competitive. It is this last feature which makes system engineering a fascinating discipline, and in addition the methods which have been developed can have far-reaching effects. We may not, therefore, be able to foresee, at the beginning of the project, the final outcome of our labours. To give a few simple examples –

> A fully automated oil refinery run by a few engineers with the help of a sizable computer is an example of a system. A satellite communication system complete with orbiting satellites, several command stations on the ground, transmitting and receiving units with associated computers is another example. A completely automatic remote-controlled power station system comprising 30 or 50 power stations connected onto a common net, covering a very large area, under the command of an engineer with the help of one or two computers, would be another example, and so on.

Furthermore, a more accurate model of a system would need to include the human operator at the various points in the system, if any accurate prediction of the behaviour of the system as a whole is to be achieved.

The question to which we address ourselves now is: Where does one start with the design of a complex system, if we have never designed a similar system and are not even in a position to comprehend it?

It is at this juncture that the new technique of modelling is of immense value. We create conceptual models and mathematical ones; we carry out thought experiments; we model simplified systems on a computer and proceed to more complicated ones. In this way we grow richer in knowledge, in understanding, and in experience, until eventually we are ready to build a prototype model to gain some more experience on the way to reaching the final objective (Fig. 1.1). The technique here is analagous to that of a child playing successively with more and more sophisticated toys, until finally he is ready to venture into the real world.

To give a specific example, if one were asked to design a motor car, then on the first attempt one would make a rather poor job of it. But the second attempt would be much better because one would be that much richer for the experience of having designed a car before. At any rate, even with the first design one could make some sort of a start. However, if the problem were the following:

> To design a car for an inhabitant on an earth-like planet in the distant world of Andromeda Nebula, no one on this earth could make any sensible start towards the project. The reason being that we would not know the geometry of the driver for whom it is intended, neither his weight nor his particular likes and dislikes, etc. In other words we have no knowledge of the environment in which the machine is to function.

1.2 A framework

If one were asked to name the most important assets which an engineer should have, then without hesitation most of us would name, apart from professional excellence, imagination and engineering intuition. Yet intuition is something difficult to define and even more difficult to impart, but usually amounts to an ability to make judicious guesses in the absence of adequate supporting evidence. There is plenty of evidence that a good judicious guess can circumvent years of expensive research.

While in all probability intuition cannot be taught, one can be helped greatly in making inspired guesses by acquiring a knowledge of the value of some key parameters having a bearing on one's decisions. This can be achieved quite effectively by listing the relevant parameters in order of their magnitude and committing to memory some of the key points on the lines indicated in the following examples.

* see for example: Systems Engineering Handbook, P.E. Machol (Ed). (McGraw Hill, 1965).

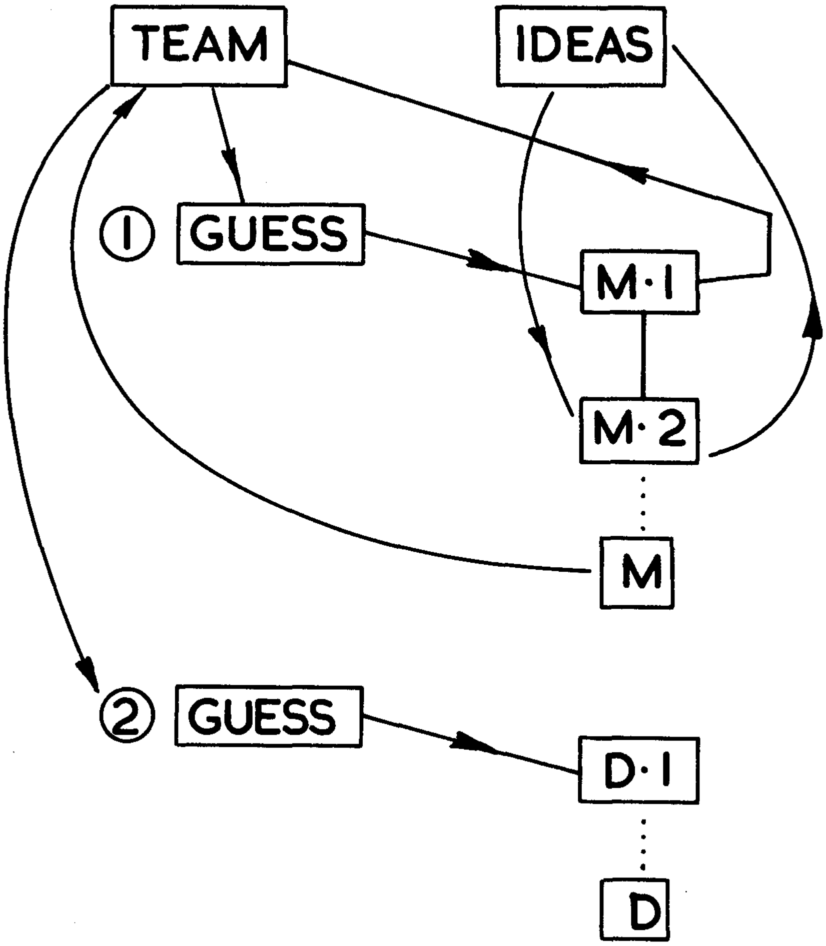


FIG.1·1. MODEL & FEED-BACK DESIGN TECHNIQUE OF A COMPLICATED SYSTEM

1.3 **Examples**

We could take as the first example the physical quantity "length". In the middle of the scale we can mark a point denoting the height of a human being (of the order of 1m) and relate other objects to it on a logarithmic scale (Fig. 1.2). In this way we can compress an enormous range of values and compare their relative magnitudes. We note that the total range of physical distances known to man is of the order of magnitude of 10^{40}, an enormous number beyond human comprehension (it is known under the name of Eddington's number).

A similar exercise performed with respect to the physical quantity time gives the chart shown in Fig. 1.3. Here too, we find an enormous range of values.

A variety of such charts can be prepared concerning various physical and other quantities. These form a most useful reference framework in which new parameters can be placed in proper perspective.

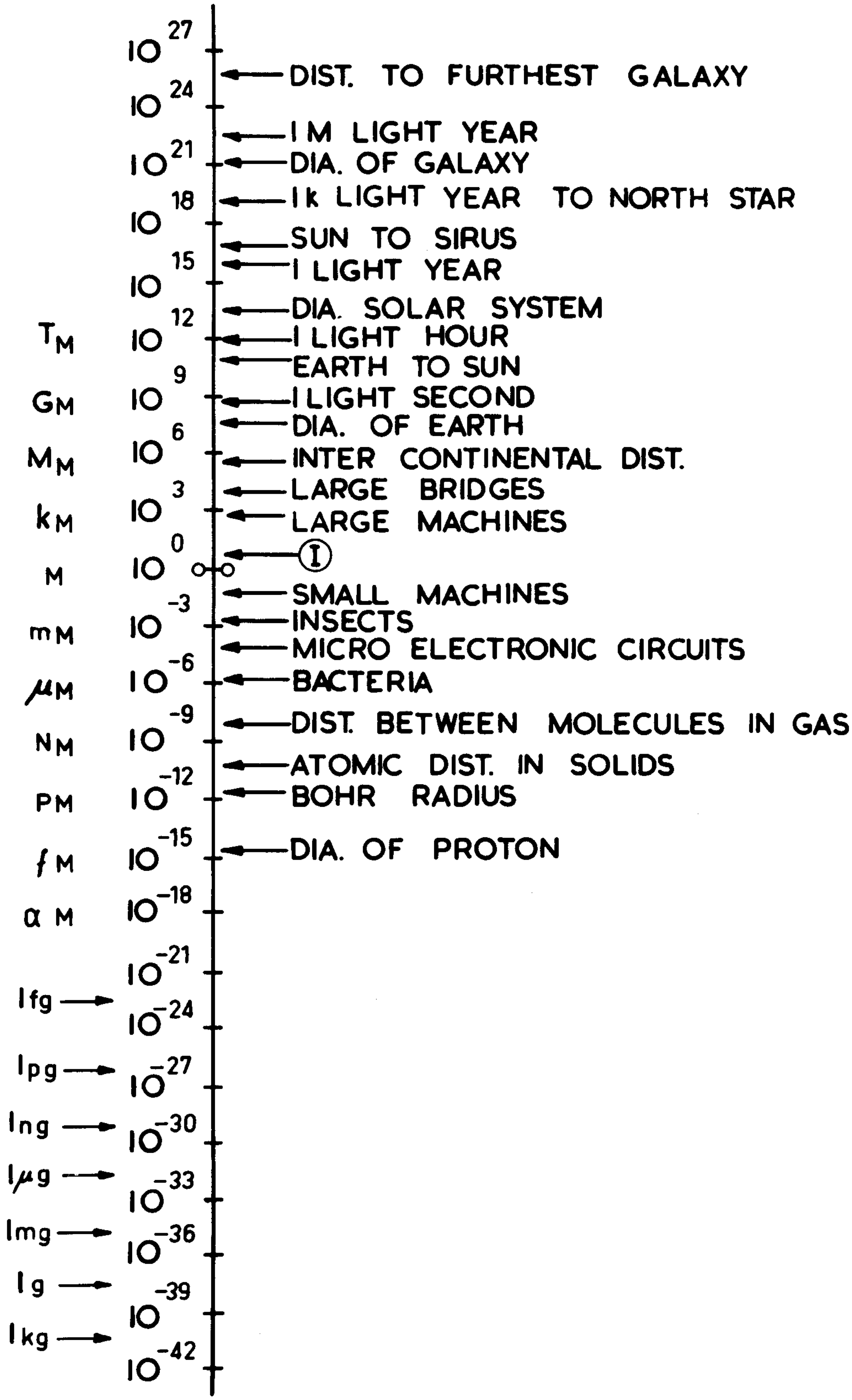


FIG. 1·2. ORDER OF MAGNITUDE OF THINGS AROUND US: LENGTH

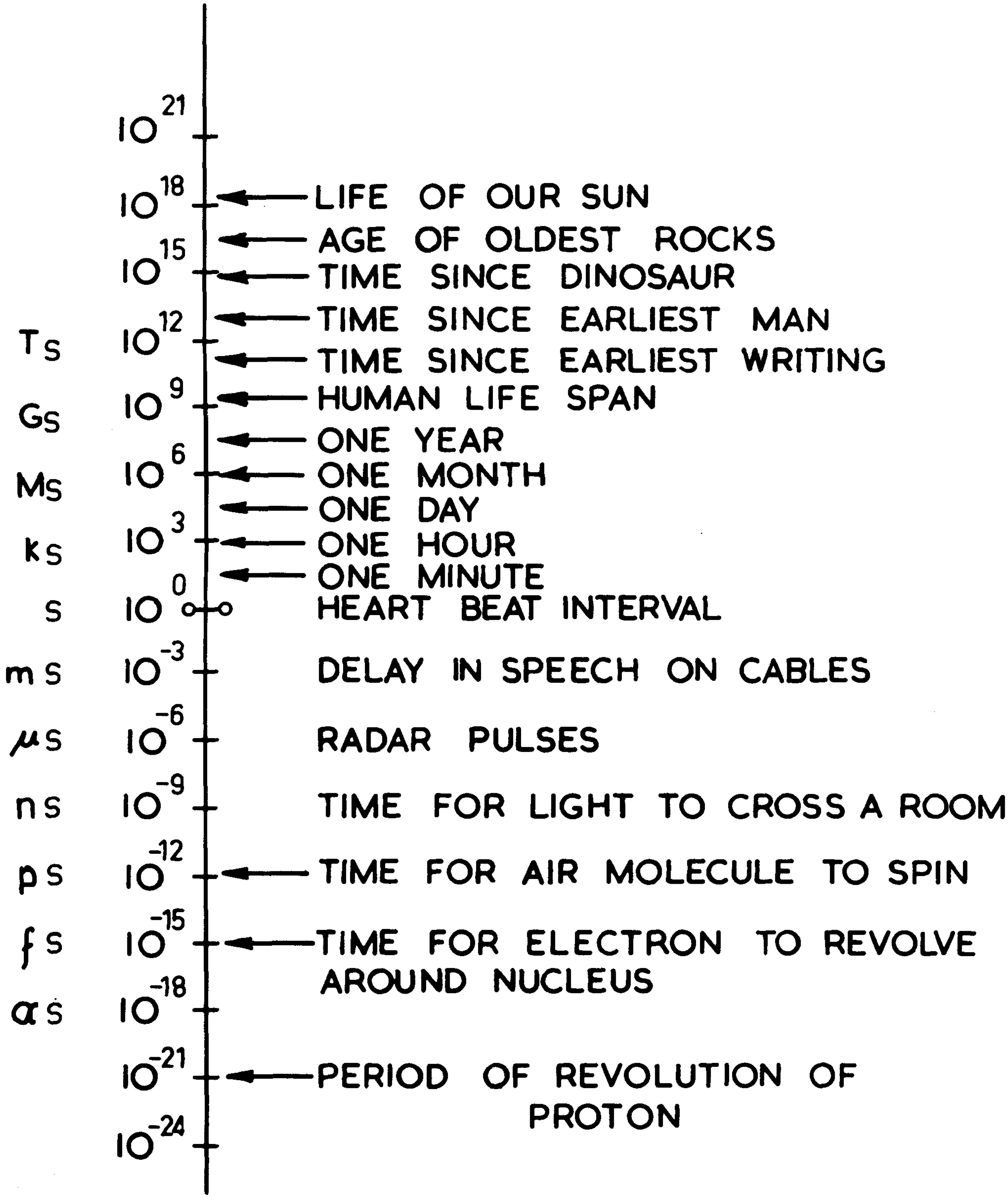


FIG. 1·3. ORDER OF MAGNITUDE OF THINGS AROUND US : TIME

1.4 Exercises

Prepare an order of magnitude chart based on

1. weight/mass
2. money and expenditure
3. any other quantity which you feel is important in Applied Science, or one which holds a fascination for you.

In 1 take 100 kg. the weight of a human being, as a starting point and work through many orders of magnitude to the mass of a galaxy and the whole of the universe at one end, and down to molecules and electrons at the other.

In 2 take your pocket money as a starting point and go through many orders of magnitude, including expenses on large scientific projects, centres of education, of second World War, or war in Vietnam, of cancer research, the national income of various countries, waste caused by hunger, inefficiency, incompetence or illwill, the value of various static assets, etc.

1.5 Suggested reading material

1. BAILEY B. and MORGAN D. "Thinking and Writing" (Rigby 1966)
2. MILLER, G.A. "Language and Communication" (McGraw Hill, 1963)
3. PHYSICAL SCIENCE STUDY COMMITTEE . . . "Physics" (D.C. Heath & Co., Boston, 1960).

CHAPTER 2

Energy, Power and Intuitive Notions of Information

"There is a grandeur in this view of life,
with its several powers........"

(From the concluding sentence
Charles Darwin"The Origin of Species")

CHAPTER 2

Energy, Power and Intuitive

Notions of Information

2.1 The Meaning of words.

We find in the Oxford English Dictionary the following primary definition of *energy:* "Force, vigor, (of speech, action, person, etc.) active operation, individual powers in exercise of authority and also ability." Energy therefore, can mean all or any one of those things.

The word *power* in turn has the following primary dictionary definitions: "Ability to do or act, a particular faculty of body or mind, vigour, energy, active property, government, influence, authority, personal ascendency, political ascendency, authorization, delegated authority; influential person, body or thing; Deity, large number or amount etc." We thus see that one word can mean many different things and frequently meaning is inferred in the context.

In literature (and every day usage), a certain degree of vagueness in the meaning of words is actually desirable in that it helps to convey the emotional and the artistic tenor of the material presented. In applied science the artistic and creative skills find an expression in the form of the system or the theory created, but vagueness in expression is an undesirable characteristic and an attempt is therefore made to develop a more precise language than that offered for every-day use.

Thus in science one *defines* energy as: *"Ability to do work"* (potential or kinetic), and power as *"rate of doing work"*. More precise definitions can be formulated with the help of mathematical symbols. Thus we define energy as "force times distance" and power as "the derivative of energy with respect to time", or "the rate of doing work". In addition we understand both quantities to be "scalars".

There are two points which need to be brought to light at this stage. The first one concerns the language of a scientist and its relation to human communication. Thus, with the specific reference to the meaning of the words energy and work, a scientist having understood the physics behind energy (and work) as pertaining to a physical system, *borrows* a word from everyday language – a word which has an *accepted* vague meaning – and ascribes to it a more precise but specific significance which to most people around him is strange. Here lies the first source of confusion: the scientist defines precisely the meaning of a word to signify one thing and his listeners will understand another thing. This of course does not help in the communication. Here an observation is in order.

The re-definition of the meaning of a vague word may help the scientist in his work, but it certainly does not help in the communication of his ideas to the layman. To the scientist, it would have been equally acceptable to have defined a new word and have given to this word the intended precise meaning, thus avoiding all subsequent confusion.

This is a somewhat simplified illustration of the kind of confusion which can arise among members of a team of specialists working together towards a major project, in that they fail to understand each other on account of language difficulty rather than technological complexity. This is the first point which concerns confusion which can arise from the use of words, having an established vague meaning, with new connotations.

The second aspect concerns the opposite effect. Here we have in mind the confusion which may arise from an attempt at a too rigorous (or excessively abstract) association. To give a specific example: We know that energy and power are defined (mathematically) as scalars. By this we mean that the quantity is adequately specified by a number, unlike a vector quantity which needs for its full description a number and direction associated with it. Now imagine an engineer being given a source of power rated 200 h.p. and not being told the direction in which the motor is revolving. Clearly, he would not be in a position to design the actual machine using the data given; he needs to know in addition the direction in which the motor is to revolve when delivering power.

These two examples are simple illustrations, and serve as a warning of the dangers impending in the excessive use of technical jargon. Frequently, it is this factor which is responsible for a technical or scientific discourse degenerating to a second rate debate on the language itself.

2.2 Energy and Power, the two fundamental quantities of applied science.

Every engineering system needs, in order to function a source of power. There needs to be a power distribution network to carry the energy to the various parts of the system. But the purpose for which the power flows in the system is twofold.

1. To power the system so as to enable it to perform the specific tasks.
2. To carry the information around the system; here, energy is the vehicle for conveying information.

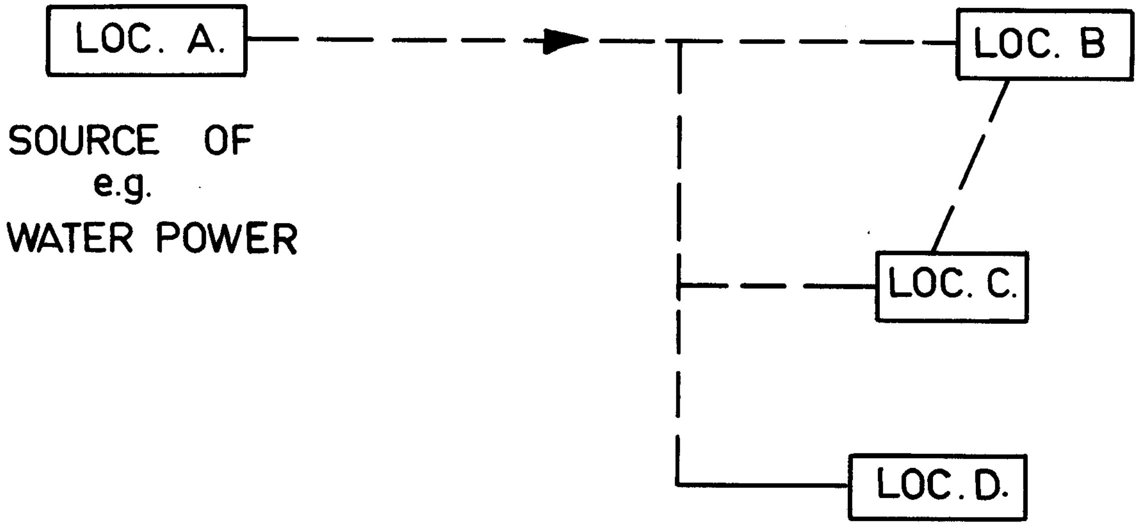


FIG. 2·1. ENERGY TRANSPORTATION

The problems associated with the first objective are those pertaining to the generation as well as transportation of energy between points distant in space (Fig. 2.1). The energy can be transported in a variety of forms such as hydraulic or electrical, or by means of compressed gas etc. But the actual manner of distributing the energy is of lesser importance in systems engineering, although usually for convenience, the energy is transported by electrical means.

Many years ago in electrical engineering we used to differentiate between power engineering (heavy current engineering) and light current engineering. The former was concerned primarily with generation and transportation as well as utilization of electrical power, while light current engineering covered such aspects as electronics, telephone, radio and so on. The reason for this subdivision was in part due to the fact that light current engineering was concerned principally with aspects of communication and control engineering, where only small amounts of power were involved. However, times have changed, and we now realise that we need energy for communication, and the larger the range and the greater the quantity of information to be conveyed, the greater the demands on power.

To give a specific illustration: If a teacher speaks to a class of some 10 or 20 students, it is then quite sufficient for him to speak at the normal volume of his voice, but if the class is increased to something like 50 or 100 students, he needs to raise his voice and perhaps, at the end of the lecture, he might feel tired. When it comes to even larger audiences of the order of many hundreds or thousands of listeners, then the speaker in order to make himself heard and to convey the information intended, needs powerful amplifiers to strengthen his voice. Thus we see that the larger distance and the greater the quantity of information to be conveyed, the greater the demands on the power. Nowadays, radio transmitters have powers typically of the order of 100 kilowatts so as to enable the transmitters to broadcast the information over a large area of influence. Similar observation would apply to television and other mass media communication.

We have progressed a long way from the early days of electrical engineering, and nowadays, we have communication systems where the transmitters have powers of thousands of kilowatts and even of the order of many megawatts in order to achieve the objectives assigned to such apparatus. So, the distinction between heavy current engineering and light current engineering can be said to have disappeared, but we still have the conceptual difference in that, in power engineering the primary concern is to transport energy between distant points in space (Fig. 2.1). Here, efficient transportation of energy is of paramount importance.

With communication systems, however, while admittedly power is being consumed to a very large extent, the primary objective is not to convey power, but to convey information. (Fig. 2.2).

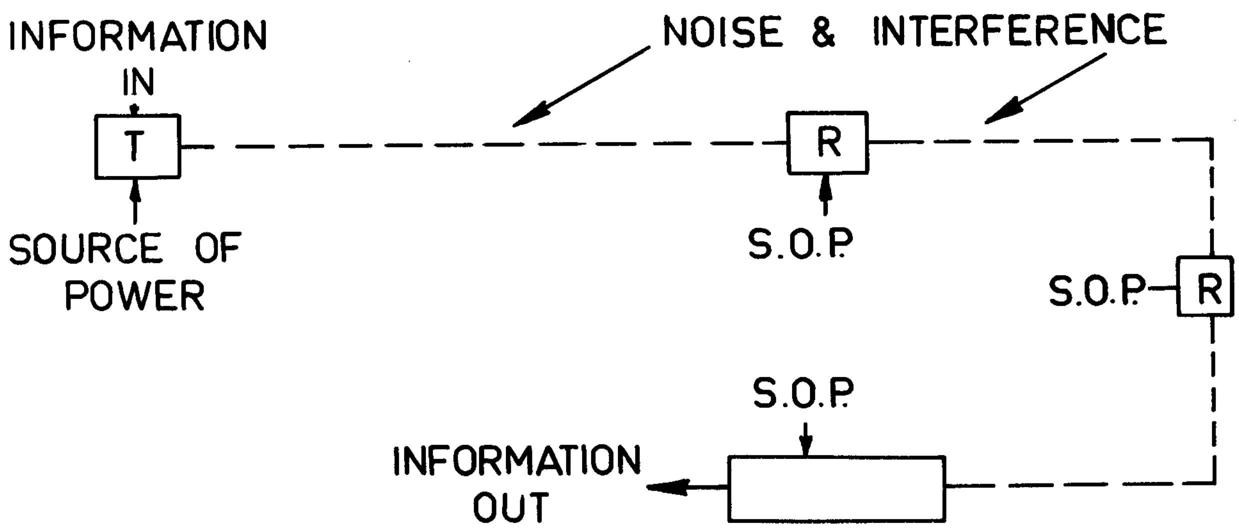


FIG 2·2 INFORMATION TRANSPORTATION

2.3 **Energy and Information.**

Having established that energy needs to be expended to convey information, it will also be appreciated that energy must be expended to gain information. If one wishes to acquire a great deal of information, then a very substantial amount of power needs to be expended. By the same token we see that energy must be expended to control energy, and the greater the amount of energy to be controlled and the greater the precision required in controlling the energy, the greater will be the need for the controlling power. From reasoning of the above nature, various corollaries follow. Thus we find that;

(1) The higher the accuracy of measurement, the greater the demands on power, and

(2) The higher the precision of control, the greater the demands of power.

Therefore, there is a limit to the accuracy and the precision with which any system can operate.

Thus we find, that there is a limit to the rate with which information can be generated, and also there exists a limit to the rate of construction of devices. The higher the rate, the greater the power needed to achieve the objectives.

The final point which needs to be made, is that every engineering system exists in an environment, and that any environment is a part of the Universe at large. The environment as such, possesses a temperature, and therefore it is subjected to Brownian motion. At the same time, the act of measuring any part of the system implies interference with the system or the thing which we set out to measure. Thus the act of communication, in part modifies the communicants as well as the environment; the act of controlling, in part modifies the thing we set out to control, etc.

The reasons behind such characteristics of physical systems are intimately related to the fundamental concepts of physics. Here we refer to the three fundamental constants of physics. These are:

h (Planck's constant), k (Boltzmann's constant), e (electronic charge).

These respectively, relate to the minimum discernable energy on account of the quantum nature of matter. These are:-

hf, h = 6.625×10^{-34} joule sec.

kT, k = 1.38×10^{-23} joule/degree K

e = 1.602×10^{-19} coulombs

All three quantities, i.e. the quantum of action, hf, the thermal energy per degree of freedom, ½ kT, and electronic charge e, are, in various forms, responsible for the existence of errors or the so called noise in engineering systems. These are the intrinsic sources of noise, but for completeness we should add the extrinsic forms of noise which relate to various errors of human and other origin.

2.4 Exercises

1. Calculate kT & hf for a range of temperature and frequency values and try to appreciate the magnitude of energies involved. Convert the energies obtained into eV.

2. Sketch a chart (starting with kT & hf at one end) of magnitude of energies and powers around us including those pertaining to large power stations, lightning discharge, space rockets, motor car, food we eat, computer elements, biological organisms, neurons in human brain, etc.

2.5 Suggested reading.

1. PHYSICAL SCIENCE STUDY COMMITTEE: "Physics" (D.C. Heath & Co., Boston, 1960)

2. BRILLOUIN, L. "Science and Information Theory" 2nd Ed. (Academic Press, 1954).

3. PIERCE, J.R. "Symbols, Signals and Noise: The Nature and Process of Communication" (Hutchinson, 1962).

CHAPTER 3

Patterns, Waveforms and Scientific Measure of Information

"....we shall exert all our energies towards the shaping of a plan ...."

(New Year message to the Australian people by Prime Minister John Curtin, 29th. December 1941.)

CHAPTER 3

Patterns, Waveforms and Scientific Measure of Information.

3.1 Concepts and definitions.

It is a common experience, that the human voice conveys information. However, speech is not the only means by which humans communicate; facial expression as well as gesticulation also help in communication. We thus see that many aspects of human communications are inherently complex.

But suppose that we restrict ourselves to communication by human voice. Clearly, words are not the only factors in communication. Intonation, volume, and even pitch of voice affect information profoundly. This being the case, how can one measure information? The simple answer is that one cannot really measure information, or communication, in the everyday sense, any more than one could measure the suffering of a grief stricken person. But this need not deter us from an attempt at measuring the seemingly impossible. By trying to measure, we improve our techniques, we gain a better understanding, and in so doing, we gradually appreciate the difficulties involved. Finally, we will find a way of defining something that relates to the quantity of interest, a quantity which can be measured.

Table 3.1 Power of Acoustic Sources

Source	Power in watts.
Orchestra of seventy-five performers (full volume)	70
Bass drum	25
Pipe organ	13
Cymbals	10
Trombone	6
Piano	0.4
Bass saxophone	0.3
Double bass	0.16
Orchestra of seventy-five performers (average loudness)	0.09
Flute	0.06
Clarinet	0.05
French horn	0.05
Bass voice	0.03
Alto voice pp	0.001
Average speech	0.000024
Violin at softest passage	0.000004

The real purpose in trying to measure infomation and communication lies in our desire to carry out quantitative comparisons between information carrying systems, but in doing so, the scale of measures must be meaningful in an engineering sense.

When dealing with voice sounds, an obvious choice would be, for example, to measure volume of the acoustic disturbance. If we did that, we would obtain results as set out in Table 3.1. We note from the entries in the Table, that the average power of human voice is of the order of 24 microwatts, which is an exceedingly small power, when compared with

power of other sources such as a symphony orchestra (of the order of 70 watts). Thus, we see, that power is an important consideration, but on its own, does not help us in speech recognition.

Another possible parameter to study would be the pitch of the voice. An even better description might be to record the actual waveforms of the voice, or the spectrograms of different sounds, etc. But when we come to examine such evidence, we quickly realise that the parameters which we are studying do not really help us in speech recognition. We therefore conclude that the measures discussed do not necessarily relate to the information bearing capacity of acoustics sounds as generated by humans. Such studies, although legitimate attempts at scientific study, produce quantitative data which is of very limited use.

The complexity associated with the subject of communication can be further illustrated by observing that much information is also conveyed by means of pictures and objects, and even here the description is imperfect in that the same number of objects in different juxtapositions will convey different information. Thus we come to appreciate the tremendous variety of ways, means, and different languages which humans have at their disposal to communicate.

3.2 Attempts at measuring information.

3.2.1 Background.

"To inform", according to the Oxford English Dictionary, is "to tell" usually an item of news and "to communicate" is "to impart something". The informer is the source of information but the process of informing does not necessarily ensure that the information has actually been received or understood. Communication, on the other hand, does imply at least a partial two-way flow of information, and some degree of understanding is also involved.

With modern systems, information is usually carried and represented by a variety of electrical signals, and these can be related to the data or patterns transmitted. For example, data on temperature and pressure relate to meteorological patterns; salinity and temperature of sea water give patterns of oceanography etc. There are patterns of ecology, of astronomy, of human behaviour, of heredity and many others, but this is of lesser concern to communication theory which deals primarily with translation of patterns (or data making up the patterns) rather than with what the patterns stand for.

There is an endless variety of different patterns but any family of patterns which can be resolved into its constituent data can be analysed in a systematic manner using information theory. This is the basis of scientific analysis and any branch of our knowledge which has reached the degree of sophistication of being able to be analysed in the above manner, can be subjected to the quantitative scrutiny of information theory. But, no amount of analysis will tell us whether the data which make up the different patterns "make sense", for this would be semantic information which is outside the domain of Information Theory.

One set of patterns can be translated on a one-to-one correspondence into another set. Such processes are known under different names of "encoding" or "modulation" depending whether we deal with digital or analogue data. It is also true that digital data can be translated into a corresponding analogue form and vice versa, and the real reason why this is possible is because any analogue pattern realisable in the physical world contains but a finite number of degrees of freedom: the number of independent data needed to describe a pattern is, therefore, always finite. (c.f. Section 4.3).

In trying to apply Information Theory to an engineering system, the first step is to find means for translating the various patterns, which we as humans perceive, into some form of quantitative language, and to do that we must devise means for measuring information. One possible way would be to take a poll of opinions and ascribe a scale of values accordingly. This is in fact widely practised in a variety of situations such as food testing, where a group of connoisseurs arrange the produce in their (individual) order of preference, numbering them 1, 2, 3, etc. Then, by taking a weighted average of the results so obtained, we arrive at a quantitative measure which relates to the quality or value of the commodity judged.

In communication engineering a similar approach is frequently used when studying such aspects of human communication as speech intelligibility. However, on the whole, the approach is difficult to generalise, and is used only when other attempts at measuring information fail.

The idea behind the thought leading to the scientific measure of information can be best explained using a simple example. Suppose the objective is to measure the aesthetic value of a lady's dress. The task as stated here, would appear to be impossible to implement. Yet a systematic and scientific approach does enable us to go a long way towards the objective. One simply invents a scale of measure which it is thought relates to the quantity of interest. In this way, the position of the hemline with respect to a convenient reference mark could provide such a scale of measure. If we then say the hemline is 20 inches above the ground, then the statement becomes quantitative and acquires scientific meaning. This is a parallel of what information theory tries to do. Indeed, it is consideration of examples of this class, of successively increasing degree of complexity that led to the development of Information Theory to the form in which we know it today.

The subject of communication was outside the domain of science until a sufficiently rigorous definition of "information" and "communication" was established. Undoubtedly, this was a great step forward, but at the same time, it

opened up a way to endless confusion (in various disguises), by accepting the words "information" and "communication" and giving them a new meaning which was strange.

Viewing this aspect in retrospect, perhaps much of the confusion could have been avoided by defining new terms such as "data extraction", "data transmission", "data processing" and the like, because in reality this is what we are concerned with. It does not really help to say that a telegraph message "arrive tomorrow" carries 2.88 bits of *information:* it would really be better, and less confusing, to say that the given telegraph message takes 2 bits of *data to transmit,* notwithstanding any meaning or information contained in it. The scientific usage of the words "communication" and "information" is now well-established and we shall therefore use these words in this newer sense rather than accept the common usage.

3.2.2 **Brief History**

Man, quite early on in history has been aware not only of the need to communicate but also of the existence of different degrees, or levels of communication. Common language was found to be a great asset, while loss of a channel of communication, such as sight, was recognised as unwelcome. In this way "language" and "channel" came to be accepted as necessary for communication. No measure of information was invented until this century, but it was accepted, from experience, that given a language and a channel we also need time to communicate. In general, the longer the available time the better the chances of conveying more information over existing channels, or of being better understood. These qualitative notions of information and communication were in existence a long time before communication theory was invented.

The need for a convenient measure of information arose from engineering problems encountered with early communication systems. It was Morse and his co-workers around 1840 who devised a nearly optimum set of symbols for the early telegraph, by assigning simple symbols to the frequently occurring letters of the alphabet and reserving the more complicated symbols for those with lesser probability of occurrence (Fig. 3.1). It is also interesting to know that the frequency of occurrence of various letters of the alphabet in the English language was deduced from an examination of fount as kept by printers. It is here that we can see an early attempt at a successful solution of optimum coding.

SYMBOL	ORIGINAL MORSE CODE	FREQUENCY OF OCCURRENCE OF PRINTERS TYPE	p (j)
E	–	12,000	.13105
T	——	9,000	.10468
A	– ——	8,000	.08151
I	– –	8,000	.06345
N	—— –	8,000	.07098
O	– –	8,000	.07995
S	– – –	8,000	.06101
H	– – – –	6,400	.05259
R	– – –	6,200	.06882
D	—— – –	4,400	.03788
L	————	4,000	.03389
U	– – ——	3,400	.02459
C	– – –	3,000	.02758
M	—— ——	3,000	.02536
F	– —— –	2,500	.02924
W	– —— ——	2,000	.01539
Y	– – – –	2,000	.01982
G	—— —— –	1,700	.01994
P	– – – – –	1,700	.01982
B	—— – – –	1,600	.01440
V	– – – ——	1,200	.00919
K	—— – ——	800	.00420
Q	– – —— –	500	.00121
J	– – – –	400	.00132
X	– —— – –	400	.00166
Z	– – – –	200	.00077

Fig. 3.1 Frequency of letters in an English text; Morse Code.

Information theory in its modern form started with the work by Nyquist (1924) and Hartley (around 1928) and was developed into a complete theory based on firm mathematical foundations by Shannon (1949). Around the same time, Wiener developed a complete theory of optimum detection of signals which differed somewhat from Shannon's approach. The two theories, in a way, could be said to complement one another, and practically all recent developments in this field are related to one or the other of the two approaches.

Much of the formalism of communication theory is submerged in mathematical exposition; this is its strength. But, at the same time, it is its weakness, in that it appears to have little in common with reality. In fact, it has been said, that a branch of science which is in its infancy needs a great deal of mathematical theory to support it, but once the necessary knowledge has been secured, then the phenomen can be explained without recourse to complicated mathematical analysis. This is partly true of Information Theory in that the formalism is essentially abstract.

3.3 Pattern generating capacity of a set: A possible basis for a scientific measure of information.

3.3.1. Patterns and Information.

Fig. 3.2a represents a simple set. The set is defined by saying that there is one site and a convention (or rule of language) that the site can be either vacant or have a black spot within it. Clearly, such a set can be made to represent at most two distinct patterns: (1) a vacant site, (2) a site with a black spot in it. We therefore say that the pattern generating capacity of the set is 2 patterns.

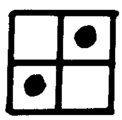
CONFIGURATION		INFORMATION CAPACITY	
		PATTERNS N	CAPACITY BITS
(a)		2	1
(b)		2	1
(c)		$2 \cdot 2 = 4$	2
(d)		$2^3 = 8$	3
(e)		$2^4 = 16$	4
(f)	q sites	2^q	q

FIG. 3·2. CAPACITY OF A SET.

Fig. 3.2b represents a different set, in that there is a different convention: Whereas previously a black spot was the alternative to a vacant site, in the present case, a site coloured completely black is the alternative to the vacant site. The two sets, therefore, differ in *form,* but are equivalent as far as pattern generating capacity is concerned.

Putting aside, for the moment, the question of form, we can extend the above concept to more complicated patterns. Thus with two sites (Fig. 3.2c) the pattern generating capacity of the set is 2 x 2 = 4 patterns. With 3 sites (Fig. 3.2d) the capacity is $2^3 = 8$ and with 4 sites it is $2^4 = 16$ patterns. In general, if there are q sites, the maximum number of distinct patterns which the set (language) can be made to represent is $2^q = N\ (q)$. It should be noted that the exponential law, giving the number of patterns in the set, is an important property, while the binary base is a direct consequence of the fact that there are two choices per site (either vacant or occupied): If there were 3 attributes per site, then obviously the base would be 3 and the law for number of patterns which the set can generate would then, of course, be 3^q; and correspondingly for other cases.

Knowing the property of the set, it would be adeuqate to describe the capacity of the set in terms of the exponent q. This has its advantages as will be apparent from the following discussion.

With reference to Fig. 3.2a, we note that one decision is needed (either *yes* the site is vacant, or *no* the site is occupied) to determine the pattern. Clearly, for a set with 2 sites, 2 decisions are needed, and for a set with q sites, q yes/no decisions need to be made to determine which particular pattern has been chosen out of the family. Thus q is the number of binary decisions (yes/no, black/white, etc.) needed to identify the particular pattern unequivocally. Therefore, q is the binary information, measured by the number of binary decisions (or bits for brevity) needed to reconstruct any one pattern of the set. In fact, q bits of information is a characteristic of the set.

There are, therefore, two ways of describing the pattern generating capacity of a set:

(a) by the actual number of distinct patterns which the set can generate

$$N = 2^q \text{ for binary patterns} \tag{3.1}$$

or

(b) in bits

$$C = q \text{ bits} \tag{3.2}$$

The utility and simplicity of the concept of measuring information in bits is the principal attraction of the method.

3.3.2. Patterns and form

Let us now return to the question of form. We observed in connection with the patterns illustrated in Fig. 3.2a and b, that the two sets, although identical as regards the pattern generating capacity, differ in form. Obviously, if form is important, then it can also be taken into accoant by a suitable digital description. Thus, with reference to (a) and (b) in Fig. 3.2, if we say that there are 3 attributes per site (vacant, black spot, completely black) then (a) and (b) will be subsets of a one-site set with three attributes. The total number of patterns which the set can be made to represent is 3 (not 4, since the vacant site repeats in both subsets). This also follows for an alternative representation by means of a Venn diagram (Fig. 3.3).

In general, if there are n attributes per site of a set of q sites, then the pattern generating capacity becomes*

$$N(h, g) = n^q = 2^{q \text{ lgg } n} \tag{3.3}$$

Therefore, the information conveyed by any one pattern of the set is q lgg n bits. Consequently, it follows, that the information can be "translated" on a one-to-one basis and be represented by a binary set of q lgg n sites.

Varying the attributes of the sites is one way of changing the form of a set. There are, however, many other alternatives. Fig. 3.4b, shows a number of sets of identical capacity but differing in configuration. No doubt, we could consider such configurations as being subsets of a much larger set, and take configurational changes into account quantitatively, but if we do not put any restrictions on the configurational changes, then the mother set will grow in size accordingly. In the end, to describe a particular pattern and the configuration may require an inordinate quantity of data, so that the description in such terms becomes impractical. Such are the difficulties associated with detailed description of form. The fact, however, remains that in principle, forms of any nature can be described quantitatively, and the scientific measure of information as described above is of direct use.

* lgg denotes $\log_2$

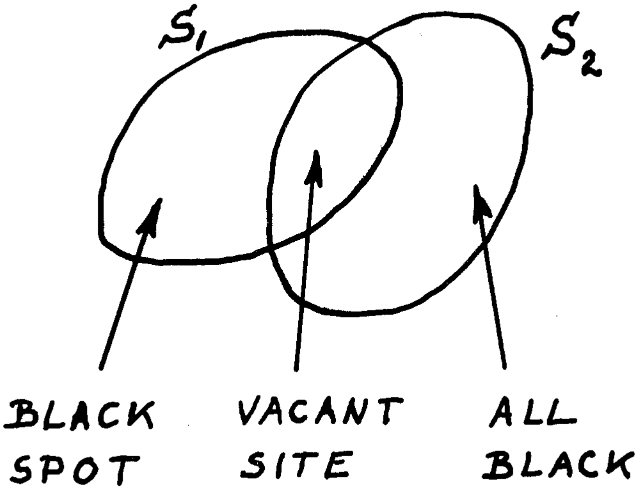


FIG. 3·3 VENN DIAGRAM REPRESENTATION.

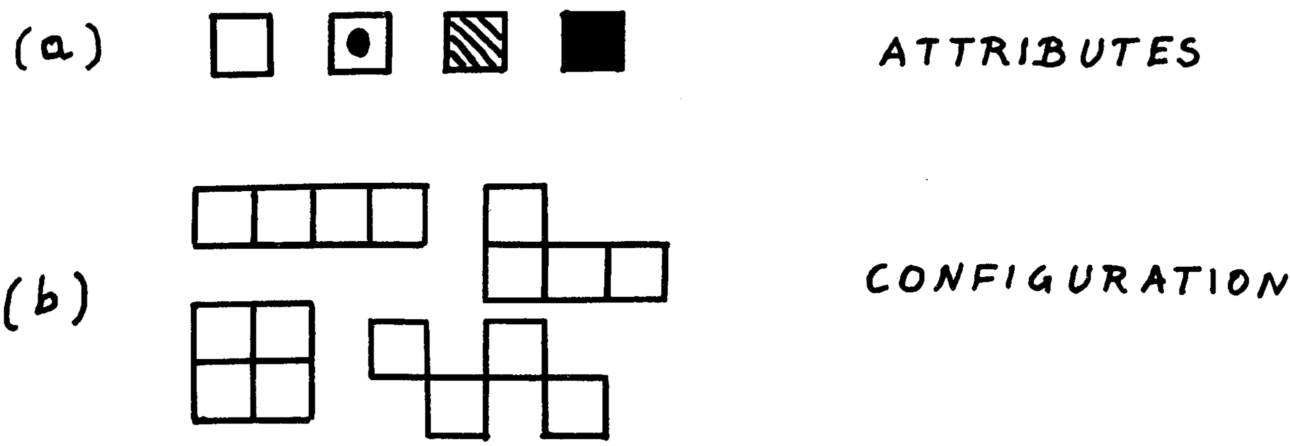


FIG. 3·4. PATTERNS AND FORM.

3.4 Waveforms; a class of patterns.

3.4.1 Patterns, constraints, redundancy.

Consider a binary set of 8 x 10 sites illustrated in Fig. 3.5a. In view of the foregoing discussion the total number of patterns which the set can generate is

$$N_o = 2^{n.k} = 2^{80} \approx 10^{24} \qquad (3.4)$$

Consequently the capacity is given simply by

$$C_o = n.k. = 80 \text{ bits} \qquad (3.5)$$

In practical applications the set just described is of a particular significance. We shall call it the checker-board pattern, and for the moment shall regard it as the mother set. We can derive from the checker-board pattern, a large number of subsets by the simple expedient of introducing constraints. But before considering constraints in any greater detail, we can remark at this stage that the checker-board pattern can be viewed as a pattern made up of a number of vertical bars with a given number of sites. Let there be n sites in each vertical bar and the number of vertical bars be k. In the example given there are 8 sites in a vertical bar and the complete pattern consists of 10 vertical bars. We can say that the information per vertical bar of the checker-board pattern in the example given is 8 bits.

There exists a large variety of ways in which constraints can be introduced into the mother set to generate different subsets. One possibility would be to restrict the number of occupied sites per vertical bar. Fig. 3.5b illustrates a particular case. Here the constraint (the grammar of the language) is the restriction that we agree to have all but two sites of every vertical bar unoccupied. The pattern generating capacity of this subset can be calculated as follows.

The number of distinct patterns which one vertical bar can be made to represent is the same as the number of ways in which two objects can be arranged among n sites, and this is clearly given by the number of combinations of two objects out of n. Therefore, the general formula is

$$N_b = \frac{n(n-1)}{2} = C_2^n = \frac{8.7}{1.2} = 28 \qquad (3.6)$$

The quantity N_b can be rewritten in the following form

$$N_b = 2^{\lg g\ 28} \qquad (3.7)$$

Therefore the capacity per vertical bar becomes

$$C_b = \lg g\ 28 = 4.8 \text{ bits} \qquad (3.8)$$

However, since the capacity of the vertical bar subset is proportional to the number of vertical bars in the set, the capacity of the set is

$$C_1 = k.\lg g\ 28 = 48 \text{ bits} \qquad (3.9)$$

The total number of patterns which the set can generate is clearly

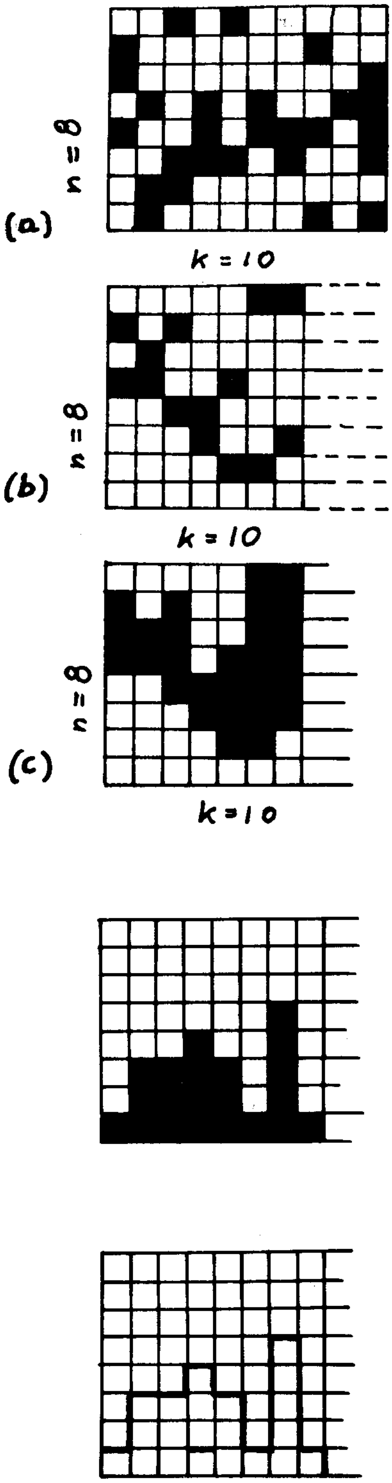
$$N_1 = 2^{C_1} = 2^{k.\lg g\ 28} \approx 2^{48} \approx 10^{15} \qquad (3.10)$$

With respect to this set, we can make a number of important observations. The first one is that the set, with the constraints discussed, has a capacity which is very significantly smaller than the mother set. The number of patterns which the constrained set can generate is smaller than that of the mother set by a factor of 10^9 or in terms of bits, it amounts to

$$C_1 = C_0 - R_1 = 80 - 32 = 48 \text{ bits} \qquad (3.11)$$

We therefore say that the constrained set is redundant with respect to the mother set by 32 bits. By this we mean, that whereas one needs 80 bits of information to determine a pattern belonging to the mother set, only 48 bits need to be transmitted to describe fully a pattern belonging to the vertical bar set having the same number of sites. The language is said to be redundant and could therefore be translated on a one-to-one correspondence into an alternative checker-board pattern having 48 sites instead of 80 sites, thereby achieving a more economical description in terms of sites.

We note, with respect to the pattern illustrated in Fig. 3.5b, that in any one vertical bar, the squares located between the occupied sites carry no information and therefore the pattern illustrated in Fig. 3.5c represents the same information as the pattern shown in Fig. 3.5b. The two sets are therefore equivalent, although their forms are different.



$N_b = 2^8 = 256$

$N = 2^{n\cdot k} = 2^{80} \approx 10^{24}$

$C = n\cdot k = 80$ bits

8 bits / column

$N_b = \frac{n(n-1)}{1\cdot 2} = C_2^n = 28$

$N = 2^{k\cdot \lg 28} \approx 2^{48} \approx 10^{15}$

$C = k\cdot \lg 28 = 48$ bits

4.8 bits / column

same

(d)

vertical bar pattern

(e)

quantised wave form

FIG. 3.5. PATTERNS CONSTRAINTS AND REDUNDANCY.

3.4.2 **Waveform: A subset of a checker-board pattern.**

Consider now a different constraint on the checker-board pattern. Let there be only one site occupied in a vertical bar. Here the total number of patterns which a vertical bar could generate is clearly, $C_1^n = 8$. The capacity per vertical bar is therefore

$$\text{lgg } n = \text{lgg } 8 = 3 \text{ bits} \tag{3.12a}$$

The capacity of the whole set is therefore

$$C_2 = k.3 = 30 \text{ bits} \tag{3.12}$$

and the number of patterns which the set can generate is given by

$$N_2 = 2^{30} \approx 10^9 \tag{3.13}$$

We note that this set has higher redundancy than the set illustrated in Fig. 3.5c, by a factor of 10^6. Or in other words the pattern generating capacity of the present set, here called the bar pattern, is smaller than that of the checker-board pattern with the same number of sites by a factor of 10^{13}. The redundancy of the bar pattern with respect to the checker-board pattern is clearly 50 bits. Therefore, the bar pattern could be translated without loss of information into a checker-board pattern having only 30 sites.

Clearly, the bar pattern has the same capacity as the set illustrated in Fig. 3.5d, where the height of the vertical bar can be made to assume one of eight possible values.

In general, the redundancy of a bar pattern with respect to a checker-board pattern is given by

$$R = \frac{2^{n.k}}{2^{k.\text{lgg } n}} = \left(\frac{2^n}{2^{\text{lgg } n}}\right)^k = \left(\frac{2^n}{n}\right)^k > 1 \text{ c patterns} \tag{3.14}$$

The set is redundant because a number of patterns contained in the checker-board set is not used. In bits we can state

$$R = k - \text{lgg } n > 0 \text{ bits} \tag{3.15}$$

This represents the information which is not conveyed by a bar pattern, and which could have been conveyed by checker-board pattern.

It will be apparent from the above discussion, that by imposing suitable constraints or inter-relations between the alternatives of a checker-board array, one can realise a set with a much larger (or smaller) number of patterns, than that which can be generated by the checker-board array. The particular constraints are important, because they define the "grammar" of the set in a unique way. Contrarywise, unless the geometry of the set is of particular significance, no reference needs to be made to any particular form of an n.k array, because the number of patterns which the set can generate is fully determined once the total number of sites together with the rules for distributing the points among the sites are given. What is of importance, is that every set possesses a finite number of distinct patterns which characterises the set. This number is called the capacity of the set, which can be calculated from the knowledge of the number of sites and the grammar (the mathematical constraints) of the set as discussed above.

When dealing with patterns arising in different engineering systems, one needs to study the properties of the physical system from which the class of patterns of relevance can be deduced. E.g. a television picture can be regarded as an array of n.k points. Thus, if we neglect the various shades of grey*, the appropriate class is that corresponding to a chequer-board pattern and the information capacity of the patterns can be calculated from the formulae given above.

To take another example, if the n.k array is formed by the signal voltage capable of assuming any one of n different values of voltage one in each of the k intervals of time, then clearly such a waveform is an embodiment of the bar pattern (Fig. 3.5e.). The capacity of the system is therefore n^k distinct patterns in k intervals of time, or simply (k.lgg n) binary decisions. Any differences which might arise are purely configurational (therefore pertaining to form), and as such, need not be considered when calculating the capacity.

There are some simple physical reasons for thinking that a waveform is an example of a bar pattern (see Fig. 3.5d & e) This can be seen from the following reasoning. The physical attribute of voltage or current is that it must be a single valued variable, and consequently in any particular instant of time it can only assume one value (1 black point per vertical column). Hence, configurationally, a waveform can be identified with a bar pattern. As such it has a pattern capacity of n^k which, as we have seen previously, needs for its full description

* See Example 3.3 at the end of the Chapter.

$$C(k) = k.\mathrm{lgg}\, n \text{ bits} \tag{3.16}$$

A voltage waveform of twice this length would require twice the number of yes/no decisions for its description since

$$C\,(2k) = 2k\ \mathrm{lgg}\ n = 2\,C(k) \text{ etc.}$$

which is a very convenient property of C(k).

For the waveform illustrated in Fig. 3.5e, which can assume eight distinct values of voltage (0, 1, 2, 3, 4, 5, 6, 7) in any one interval 0.1 second long, the information carrying capacity of the waveform follows from the formula already given (3.12a).

Thus, it appears that the chosen scale, of measuring information in bits, appears to be a convenient one for a large class of physical phenomena. This applies to communication systems in general and the type of signals (patterns) carried by them in particular.

Thus, for the waveform shown in Fig. 3.5d we can conveniently write

$$C(10) = 10\ \mathrm{lgg}\ 8 = 30 \text{ bits/sec}$$

$$\text{or} \tag{3.17}$$

$$C(t, k) = 30t \text{ bits in t seconds}$$

which is a way of describing system capacity for producing distinct patterns. The total number of distinct patterns is 2^{30t}, or in general $2^{C(t, k)}$, and consequently to determine any one of the waveforms, one needs to make decisions (yes/no, black/white etc.) at the rate of 30 per second.

If any one pattern were selected at *random* then, clearly the chance of making a selection in any one of the one second intervals is equal to the reciprocal of the total number of patterns i.e. 2^{-30}. Therefore the figure of 30 bits describes the uncertainty as to which pattern might have been transmitted, and at the same time gives us a quantitative information as to how many bits are needed to determine which particular selection has been made. The fact that by chance we could have guessed the right pattern after only a few guesses is irrelevant, because on other occasions we would not be so fortunate, and *on average,* we could not be sure that a correct choice has been made unless we are given 30 bits of information. Thus 30 bits of information are necessary and sufficient to determine any one pattern.

We have thus established, both mathematically and conceptually, that for a waveform which can take on any one of n discrete steps in amplitude in every interval 1/k seconds long, the potential capacity is given by

$$C(n, k) = k\ \mathrm{lgg}\ n \text{ bits/sec} \tag{3.18}$$

The total number of distinct patterns which a waveform could be made to represent increases exponentially with time, and is given by

$$N(n, k) = n^k = 2^{C(n, k)} = 2^{k\ \mathrm{lgg}\ n} \tag{3.19}$$

The exponential law for the information capacity is an important property of waveforms as well as sequences, and we shall return to this point later on.

The use of logarithm to base 2 for the description of system capacity is not essential although it is a convenient choice when dealing with patterns with binary attributes per site as we have discussed above. And it is particularly convenient in connection with checker-board patterns in that the number of bits needed to describe the pattern equals simply the number of sites. In other applications, logarithms to the base 10 or to the base e are used. There are three units of information in current use. These are

1 bit being a choice between 2 equi-probable events

1 nit being a choice between e equi-probable events (equals 1.44 bits)

1 hartley being a choice between 10 equi-probable events (3.32 bits).

These units involve logarithms to the base of 2, e and 10 respectively.

3.5 Exercises

1. Calculate the pattern generating capacity of a (true) die in bits per trial. (Ans. 2.58 bits)

2. A source of signal in the form of a waveform can assume 16 distinct values in successive microsecond intervals of time independently of all the other values. Calculate the capacity of the source in bits/sec and estimate the total number of distinct patterns which could be generated in 1 second. (Ans.: roughly a number with one million significant digits which corresponds to 4 million bits)

3. A chess-board consists of 8 x 8 sites, calculate the capacity of the set in bits and then estimate the capacity of the set in terms of the total number of patterns N. (Ans. N 2×10^{19})

Assuming that a human brain contains 10^{10} effective neurons and that each one could store one pattern, compare this number with N.

If a human being could learn to memorise the pattern (N) at a rate of one per second compare the time needed to memorise all N with human life span.

Postulate an ultra-high speed machine that could classify patterns at the rate of 10^9 per second and estimate the time needed for the machine to accomplish the task.

3.6 Suggested reading material

PIERCE, J.R. "Symbols, Signals and Noise: The Nature and Progress of Communication" (Hutchinson, 1962)

SINGH, J. "Great ideas in Information Theory, Language and Cybernetics" (Dover, 1966)

MILLER, G.A. "Language & Communication" (McGraw Hill, 1963)

KARBOWIAK, A.E. "Theory of Communication" (Oliver & Boyd, 1969)

WOOD, A. "The Physics of Music" (University Paperbacks, Mathuen, 1964)

CHAPTER 4

SEQUENCES OF SYMBOLS, LANGUAGES, DATA

"When you have won, you will have everything."
(From a speech attributed by the historian Tacitus to Suetonius, the Roman Military Governor of Britain. A.D. 61)

CHAPTER 4

Sequences of Symbols, Languages, Data.

With reference to Figures 3.5d and 3.5e, it was noted that the bar pattern has a pattern generating capacity of only 30 bits, whereas the mother set, i.e. the checker-board set, has a capacity of 80 bits. To identify which particular bar pattern has been transmitted, one would not take yes/no decisions with respect to each site, as it was done for checker-board pattern, since this would be inefficient, i.e. requiring as many as 80 bits every second. Instead one develops a different strategy in pattern identification. With 8 site bar patterns, one needs to take only 3 decisions for every bar. The 3 decisions to be taken are as follows:

1. Is the height of the pattern more than half or less than the half ?
2. In which quarter is it? In the upper or the lower quarter of the half chosen in the previous decision?
3. Which eighth of the particular quarter is it?

Thus for the third vertical bar in Fig. 3.5d, the identification procedure would be as follows:

Writing Y for yes and N for no, we have the following scheme.

1.	Is the height more or less than the one half	N
2.	Is the height more or less than the quarter	Y
3.	Is it more than three eighths	N

The three answers taken together, can be put as N Y N which is the binary description of that particular bar.

Constraint can be introduced into a language in a systematic manner to increase the reliability of communication. Let us introduce an additional constraint on the bar pattern consisting of the convention that we shall make use of only levels 2, 4, 6 and 8, and suppose that we transmit such patterns at a rate of 10 vertical bars per second. Suppose also that the channel, due to imperfections (noise), can add or subtract from the transmitted height of a bar, one unit.

Clearly, the set just describedhas an *error detecting property* built into it. Thus the receiver can decide whether the message received is correct or in error. This clearly follows from the property of the set in that if the received level is 2, 4, 6 or 8 units then the receiver would decide that the message has been correctly received. But, if an odd level were received, then the receiver would immediately decide that the received character is in error, and would request a retransmission of the message.

The essence of a language having the error-detecting property is to choose symbols in such a way that the effect of an error results in the conversion of the symbol from one belonging to the set to one that does not belong to the set, and it is on this basis that the receiver is able to make a decision whether the message received is correct or in error, and take action accordingly.

The capacity of the set illustrated in Fig. 4.1, is clearly

$$C = 10 \quad \lg g4 = 20 \text{ bits/sec} \qquad (4.1)$$

or in terms of total number of patterns this is

$$N = 2^{20} \approx 10^6 \qquad (4.2)$$

It should be observed that the error-detecting property has been purchased at the expense of added redundancy. In the given example, it amounts to 10 bits per second. Therefore the pattern generating capacity of the language has been reduced from 10^9 to 10^6 but we now can communicate in the presence of imperfections in the communication channel.

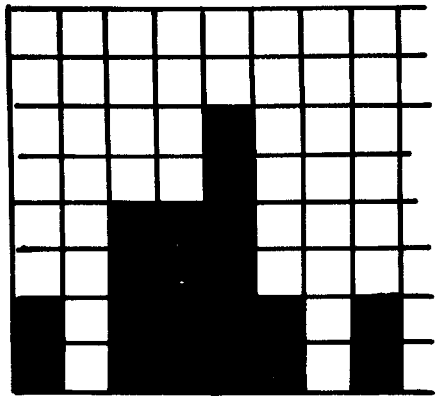


FIG. 4.1. ERROR DETECTING SET.

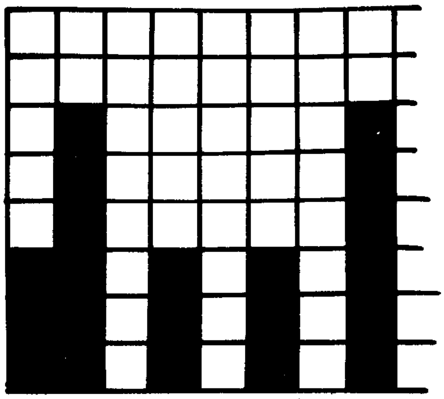


FIG. 4.2. ERROR CORRECTING SET.

Another possibility would be to agree to transmit only the levels 1, 4 and 7. Then, in spite of an error picked up in transmission, the receiver would be able to extract from the knowledge of the characteristics of the noise the correct message. The reason being that noise can at most add or subtract one unit to a vertical bar, while the levels used for communication are separated by three units. This is an example of a crude *error-correcting* language.

Most languages as we know them today, possess error-correcting or error-detecting properties to a certain degree. Thus, for example, if the message received were

LECTURES ON SYSTFM THEOPY

Then on receiving such a message, a human being would immediately guess the message intended, which is

LECTURES ON SYSTEM THEORY

This illustrates the error-detecting property of the English language.

4.2 Some characteristics of English language.

Using measures of information devised on the lines discussed above, it is possible to assess various characteristics of living languages. Thus, it was calculated if we encode English text letter by letter taking into account the frequencies of the occurrence of various letters (see Table 3.1), but neglecting the conditional probabilities, that on average we need 4.76 bits per character (this includes space as a character), but that only about 4 bits are needed if conditional probabilities are considered.

On this basis a machine can be made to produce "English words". It has been found that words so generated *resemble* English words but many are not to be found in the dictionary. Thus, some characteristics of spoken words have statistical traits but this is only part of the story.

If we encode English text by words, that is taking account of the relative frequencies of different words, about 2 bits per letter are required, and even this figure is reduced, if higher order approximations to the language are performed. Ultimately, the entropy of English text can be shown to be less than 1 bit per character. At this stage an "English text" produced by a machine resembles quite strongly written English, including good grammar at places, but lacking cohesion and sense. The text so produced wanders aimlessly and amounts to utterances with no story attached to it: perhaps the clue to semantic notions of information is to be found here.

On evidence available we can venture the following "conclusion". The statistical properties of a living language relate to one aspect but others relating to the long-range inter-relations, are left out of account. In this way the statistical properties (1 bit per character) are shared by most books yet each book tells a different story betrayed by its "phase sequence", the "semantic information". Again 4 KHz bandwidth is needed to transmit human voice irrespective of whether it carries words of wisdom or an utterance of a senseless string of words. It thus appears that one aspect of language can be described in statistical terms; its entropy.

However, language is but a carrier for human thoughts, which are conveyed by long-range interactions and juxtapositions within the framework of the language and these properties are of non-stationary nature, the semantic features. From this point of view the machine-environment interaction would appear to be of a different character to that of human-environment interaction.

Deliberations of a similar nature apply with equal force to any scientific investigation, in particular to a measuring system and the flow of data therefrom.

We shall return to this particular aspect and shall examine it in a somewhat different way in Chapter 17.

4.3 Discrete and continuous sources.

All languages which we have discussed so far belong to the class of finite languages. Such languages are made up of a finite number of symbols (patterns). E.g. with the checker-board set illustrated in Fig. 3.5a there are 256 (2^8) distinct patterns which can be accommodated in any one vertical bar. But there are only 8 distinct patterns which can be generated in one vertical bar of the set illustrated in Fig. 3.5d. In fact the capacity of the language illustrated in Fig. 3.5d is only 30 bits per second. The pattern could therefore be *encoded* into an equivalent checker-board pattern of three times ten sites in 1 second. This clearly must be the case because a checker-board pattern of 3 x 10 sites has the capacity of 30 bits per sec. It is therefore possible to translate on a one-to-one correspondence any message represented by the vertical bar set of 8 x 10 into a checker-board set of 3 x 10, without loss of information. This is a general property of all discrete languages whereby any message in one language can be translated into another language (also discrete in nature) on a one-to-one correspondence without loss of information provided the information carrying capacity of the two languages is the same. Contrarywise it is not possible to translate on a one-to-one basis from a language of higher capacity into one of lower capacity. Any attempt to do so would result in loss of information, because of necessity some many-to-one translations would have to be included.

A wave form like that illustrated in Fig. 4.3 is a pattern belonging to a different class. By an analogue signal, we understand a waveform which can assume in any particular interval of time, *any* of the values between specified limits. Thus, suppose the waveform represents an electric current as a function of time, then without further qualification, it is implied that the current can assume any of the values between certain limits, and therefore, at any particular time, the current variable can assume an *infinity* of distinct values. It therefore follows, that in a given interval of time and a given interval in amplitude an analogue waveform can, in principle, be made to represent an infinity of different patterns. As such it is an example of a continuous signal.

A source generating a continuous signal is known under the name of a continuous source. A continuous source therefore, can generate an infinite number of distinct symbols. This brings with it a mathematical difficulty in so far as we

have defined a language as being made up of a *finite* set of symbols and rules for using them. Contrarywise, with a waveform, we have a source which can generate an uncountable set of patterns and it therefore cannot be used for communication in the usual sense. So much for abstract notions concerning analogue waveforms.

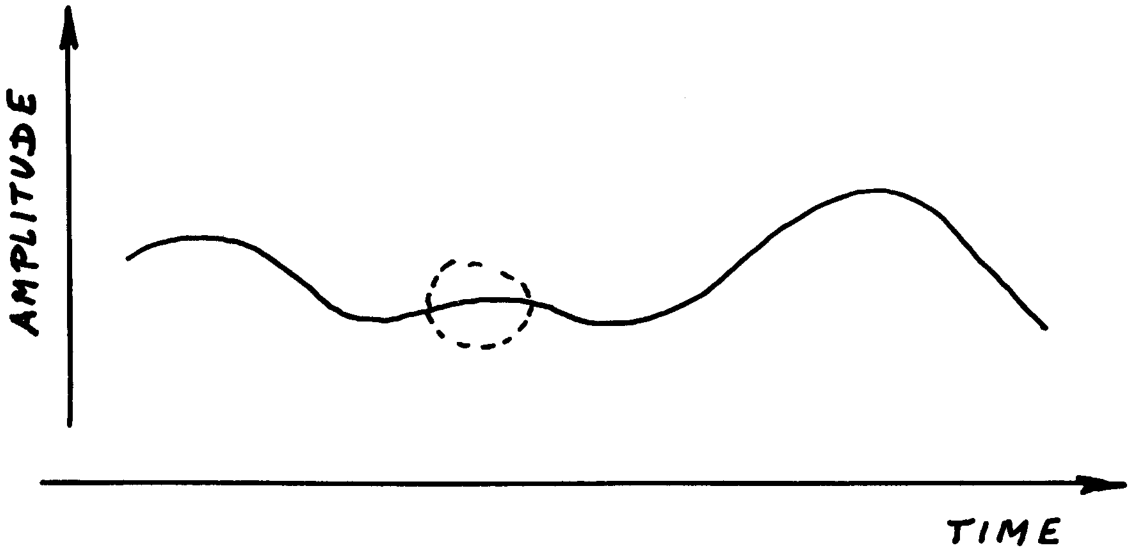


FIG. 4·3. AN ANALOGUE WAVE FORM

In practical applications, we are normally concerned with waveforms as generated by physical apparatus. Physical apparatus, systems, as well as channels, are characterised by various imperfections. There are in fact definite limits on what can be achieved. In particular, we have two factors in mind.

1. A physical system, be it a machine or a communication system, is not isolated from the rest of the world. It is therefore subject to random fluctuations (noise) which have their origin in the atomistic nature of matter.

2. Physical systems possess energy storing elements such as inductances and capacitances. The state of a physical system cannot therefore be altered significantly in an arbitrarily small interval of time.

These factors impose a limit on the amount of information which can be sent in a given time through continuous channels. There is thus

1. A minimum detectable amplitude, or energy change $\delta E = \frac{E}{n}$ (noise limitations).

2. A minimum time ($\frac{1}{K}$ secs.) which is required to make a detectable change in the state of the system (bandwidth limitations, or inertial effects).

For these reasons, a continuous channel can only *resolve* a finite number of amplitude steps in a given interval of time.

The capacity of the system, therefore, is finite. This can be further explained with the help of Fig. 4.4. Due to the inertial effects, the waveform is highly constrained and it is not therefore free to describe an infinity of different patterns. If we scrutinize an analogue waveform in minute detail as shown in Fig. 4.4 it will be noted that the voltage values in successive intervals of time differ, but very little. This, therefore severely limits the ability of the analogue waveform for generating distinct patterns. Moreover, due to the Brownian motion (factor kT) the equipment will not be able to resolve with accuracy voltage values which differ by less than a certain minimum value. It can be shown that the waveform constrained in the above described manner can be made to represent but a finite number of distinct shapes. Thus, we draw an important conclusion that an analogue waveform contains a finite amount of independent data. It is for this reason that every analogue waveform can be, by suitable processing, translated into an equivalent digital form without loss of information.

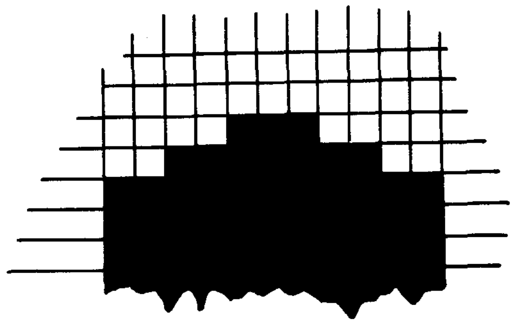


FIG. 4.4. FRAGMENT OF WAVEFORM SHOWN IN FIG. 4.3.

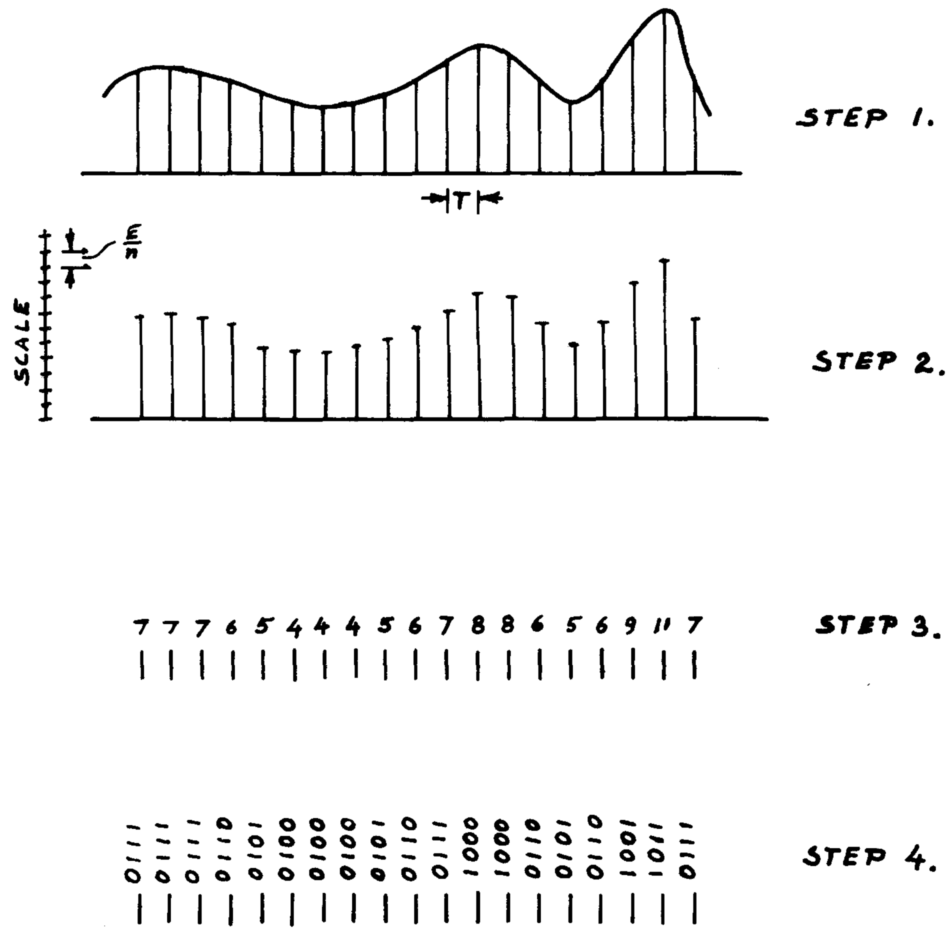


FIG. 4.5. STEPS IN CONVERSION FROM ANALOGUE TO DIGITAL FORM.

The technique for converting an analogue signal into a digital form is illustrated in Fig. 4.5. There are four steps in the process.

1. From consideration of the inertial factors (bandwidth limitation) the shortest correlation distance for an analogue waveform of the system is deduced. Let this be T. We then sample instantaneously the waveform values at time intervals which are shorter than the quantity T. In so doing, the analogue waveform becomes converted into a set of values erected at intervals of T.

2. From considerations of noise in the system we calculate the smallest amplitude step which the system can resolve.

3. We then measure the sampled values in terms of those steps. In this way, we convert the sampled values into a sequence of digits as shown in Fig. 4.5 .

4. The digital values so obtained are subsequently processed into a digital form convenient for use in the system, such as a binary language.

4.4 Languages, concept of entropy.

As explained above, language is a set of symbols (letters, characters etc.) making up a set, together with rules for using them (the grammar of the language), e.g. a binary set consists of two symbols [A, B] and the grammar might be that the symbols occur with equal probability and that there are no intersymbol constraints. Typically we would have a sequence like

$$\ldots \text{ABBAAABABBABBBBABBAAA} \ldots \qquad (4.3)$$

Clearly with such a sequence we cannot tell what the next member in the series is going to be. It could equally well be an A as a B. In fact the information needed to determine whether it is going to be A or B is just one bit. Therefore, the information per symbol of the sequence is one bit, being a characteristic of the language described.

On the other hand, a sequence

$$\ldots \text{ABABABABABABABABABABABABABAB} \ldots \qquad (4.4)$$

carries no information because from the knowledge of the pattern we can say with certainty that the symbols A and B will occur alternately. The information of the sequence in terms of bits per symbol is zero. (Two bits for the infinity of symbols)

Let us now consider a language consisting of n equiprobable symbols. From the earlier discussion on patterns, we know that the pattern generating capacity of the language is such that the information per symbol is

$$H_1 = \operatorname{lgg} n = -\operatorname{lgg} p \qquad (4.5)$$

$$p = \frac{1}{n} \qquad (4.6)$$

In the above $p = \frac{1}{n}$ – is the probability of occurrence of any one of the symbols. H_1 is the information carried by any one symbol of the language described.

To consider a somewhat more complicated example, let us examine a language consisting of n symbols with unequal probabilities.

With the symbols $x_1, x_2, x_3, \ldots\ldots\ldots\ldots x_i \ldots\ldots x_k$

we associate the respective probabilities

$$P_1, P_2, P_3, \ldots\ldots\ldots\ldots P_i \ldots\ldots P_k$$

where

$$\sum_{i=1}^{K} P_i = 1 \qquad (4.7)$$

the information carried by the x_i symbol is, by previous reasoning,

$$-\operatorname{lgg} P_i \qquad (4.8)$$

Thus in a long sequence (M large) the total information conveyed by x_i will be

$$H_{iT} = -Mp_i \; \lg g \; P_i \tag{4.9}$$

since the symbol x_i will have occurred Mp_i times.

Similar expressions would be obtained for other symbols so that the total information becomes

$$C = H_T = \sum_{i=1}^{K} H_{iT} = -\sum_{i=1}^{K} Mp_i \; \lg g \, P_i$$

The average information per symbol of the language (assuming symbols to be independent) is

$$H = \lim_{M \to \infty} \left(\frac{H_T}{M} \right) = -\sum_{i=1}^{K} P_i \; \lg g \; P_i \tag{4.10}$$

The average information per symbol so defined is known as the entropy of the language and if logarithm to the base of 2 is used as in Equation 4.10, then the entropy is measured in bits per symbol.

The average number of patterns which a long sequence can generate is therefore given by

$$N = 2^C = 2^{M.H} \tag{4.11}$$

which shows that the number of patterns which a sequence can generate is an exponential function of the length of the sequence (c.f. Chapter 3).

Strictly speaking, in all such considerations, sequences of infinite length should be considered. This is the reason for using formally the limiting operation in (4.10).

4.5 Experiments

(1) Toss a coin (say 100 times) and record your results as a sequence of +1 (stands for heads) and -1 (stands for tails). View the result so obtained as a stochastic process and discuss its characteristics.

Plot your results in the form of a function S_k (i) = +1 -1 +1 +1 -1 .... being the summation of the outcome of the original random process (the first integral). Then view S_k (i) as a random function and compute the various statistical averages and discuss fluctuation phenomena due to the finite size of the samples.

Then discuss problems relating to prediction, and ask questions of the nature: Is it possible tp predict from the past record whether the next event will be +1 or -1? Is it possible to predict from the knowledge of S_k (j) the value of S_k (j + 1) or S_k (j + 2) etc?

Finally, questions of the form: If the coin were biased how could this *information* be obtained? If, instead of tossing the coin, one were to write down the "result of experiment" by reference to tables of random numbers (say +1 for even numbers and -1 for odd ones) could this be inferred from the examination of the "experimental" results? If, instead of tossing the coin, one were to write down the "result of experiment" by reference to a column of numbers picked at random out of a book in the library (the table of numbers could pertain for example to an exponential function) could this be inferred from an examination of the results? etc.

(2) Draw a checker-board (10 x 10 sites) pattern of the type discussed in Section (3.3) by scanning the pattern row by row, write down a sequence +1, -1, -1, +1 .... depending whether the particular square is black or white. Discuss the results. Ask and try to answer questions of the type: Could the sequences obtained in Experiment 2 resemble the sequences obtained in Experiment 1? Can a pattern be identified? Need the coding method (scanning convention) be stated to encode and decode the pattern? Can the presence of a pattern be detected? Are there any "simple" or "complex" patterns? Can the sequences be more efficiently encoded? How does one define classes of patterns? etc.

4.6 Suggested Reading Material

1. MORGAN D. "Thinking and Writing" (Rigby 1966)
2. CHERRY C. "On Human Communication: A Review, A Survey and a Criticism" (Wiley, 1957)

3. KARBOWIAK A.E. "Theory of Communication" (Oliver & Boyd, 1969)

4. MILLER G.A. "Language & Communication" (McGraw Hill, 1963)

5. PIERCE J.R. "Symbols, Signals and Noise: The Nature and Process of Communication" (Hutchinson, 1962)

Chapter 5

THE SEMICONDUCTOR REVOLUTION

There was reason for fear that, like Saturn, the
Revolution might devour each of its children in turn.
Pierre Vergniaud.

Chapter 5

THE SEMICONDUCTOR REVOLUTION

5.1 Introduction

The reader may feel that use of the word 'revolution' is a little inappropriate in a scientific or technological context. In this Chapter a story of complete and fundamental reconstruction is unfolded, which has indeed the attributes of a revolution – although no actual violence has been done other than to the outlook of engineers concerned with the design of more reliable electronic equipment, of ever increasing complexity and at a cost which must always be economically feasible. As will be seen in the following sections, the revolution involved the overthrow of vacuum tubes (valves) and discrete circuit components by integrated circuits, in which a number of transistors, diodes, resistors and capacitors are fabricated together, and inter-connected, on a simgle small crystal of silicon. In order that the important engineering consequences of this new micro-electronics may be more fully appreciated, some of the essential background understanding of semiconductor and device physics is elaborated in more detail in Chapters 6–8.

5.2 Components and devices in systems

In previous Chapters of this book we have been concerned with information, and with communication, with a view to the application of these concepts to systems and computers. Systems are large, complex, man-made equipments which are designed to perform specific functions. Examples of systems include electrical power generation and distribution systems, communication systems, co-ordinated traffic lights, automated rolling mills in steel fabrication, and so on. A system is made up of a number of parts, each of which performs a specific function. We are particularly concerned in this book with those of the sub-systems which perform an electronic function such as sensing, amplification, computation, transmission, etc. Carrying this generic subdivision a stage further, each electronic sensor, amplifier etc. is found to be composed of electronic components which have specifically designed electrical characteristics.

As an example of this more detailed subdivision, consider the case of a television receiver. A receiver is one of many similar elements of a television broadcast system, shown schematically in Fig. 5.1; other parts of the system are the studio in which the programme originates, the television camera and microphone, the transmitter, and the transmitting and receiving aerials. Each receiver, as presently manufactured, is made up of a number of components contained within its cabinet: valves or transistors, capacitors, resistors and inductors, a loudspeaker, and a picture tube which acts as a transducer to convert the amplified output of the receiver into a visible pattern on its face.

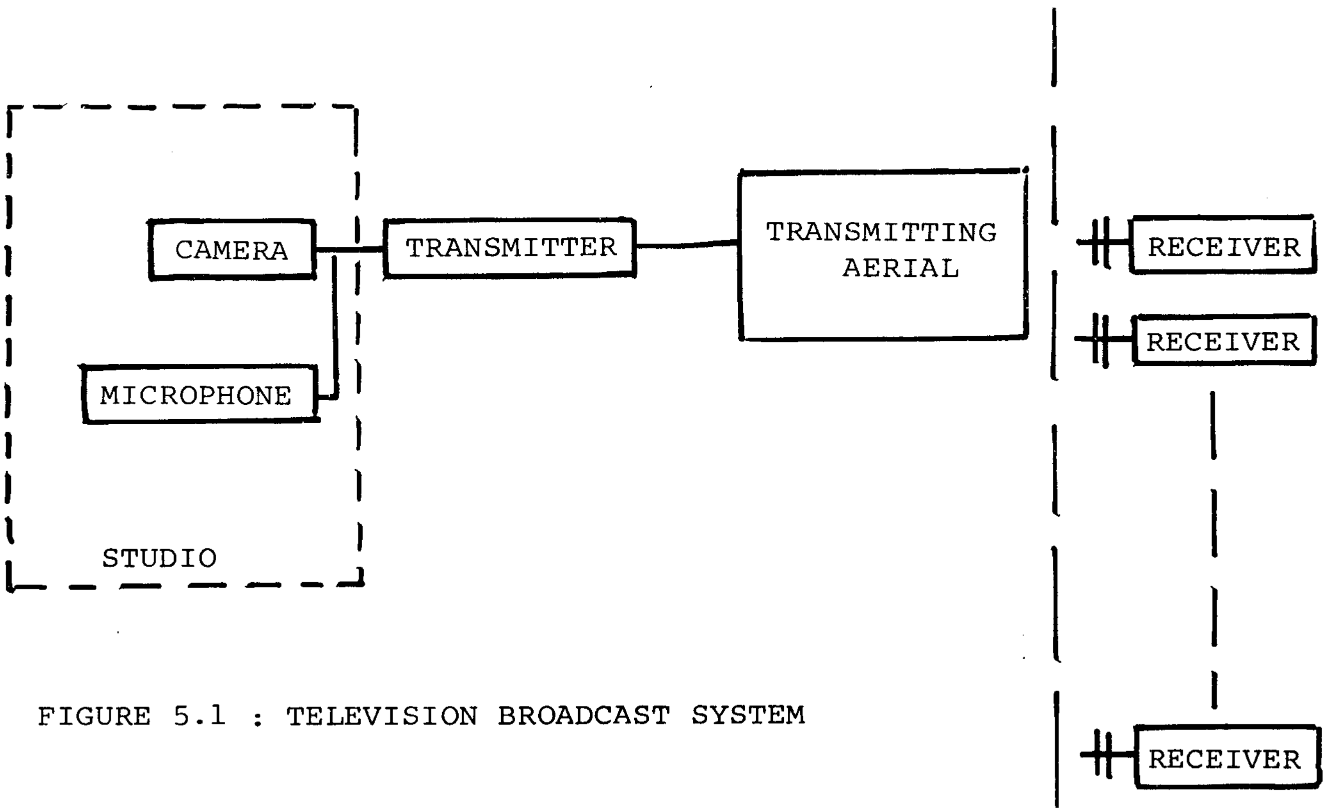


FIGURE 5.1 : TELEVISION BROADCAST SYSTEM

Corresponding to the subdivision of a complete television broadcast system outlined above, creative engineering activity has been required at three distinct levels. In the first place there was the invention and development of electronic techniques for the conversion of video and sound information into a form compatible with the known methods of transmission and reception of electromagnetic waves; in other words, the communication system was devised. Secondly, there has been the continuing task of designing each sub-system, of which the receiver is one example, so as to incorporate improved performance and reliability. An additional requirement is, of course, that such improved designs shall always be economically competitive.

Finally, we must not forget that the electronic components of which the television receiver and other sub-systems are constructed are themselves man-made: engineered from materials such as conductors, semiconductors, insulators, phosphors, and so on. The application of new physical phenomena and processes to the invention and development of useful electronic devices or components is a continuing process, and it is with this aspect of electrical engineering that we shall be concerned in Chapters 5 - 8. The development of an improved electronic component in general has some influence on the second of the levels of engineering referred to above, but we shall see that in the case of the new semiconductor components this influence has indeed been revolutionary. Semiconductor devices have even opened up possibilities of completely new systems, as will be seen in the next Section

5.3 Vacuum tubes and transistors

Of the electronic components referred to above, let us focus our attention on the active devices, those that are capable of amplification and switching. The earliest active electronic devices, the vacuum diode and triode, depend for their operation on two basic phenomena: the thermionic emission of electrons from a heated cathode, and the laws of motion of electrons in a vacuum when subjected to electric fields. A vacuum diode has the properties of a nearly ideal rectifier because the electrons emitted from the cathode will flow through the vacuum space to the anode, and thence into the external circuit, only when the anode is made positive with respect to the cathode. If a control grid is placed in the space between cathode and anode, as in a vacuum triode, the application of a negative voltage to the grid controls the flow of electrons to the positive anode, and may even reduce it to zero. Again the ability to control the magnitude of the anode current arises from the effect on electrons of the electric field set up in the vacuum space. Grid control of electron flow to the anode is so strong that one can obtain in the anode circuit an amplified version of a signal voltage applied to the grid.

The physical phenomena which form the basis of operation of transistors, and of semiconductor p-n junction rectifiers, will be discussed in more detail in Chapter 7. Suffice it to say at this stage that a transistor owes its amplifying and switching properties to the motion of electrons in a semiconductor crystal, into which impurity atoms have been introduced in a controlled way. Both the vacuum triode and the transistor are three-electrode devices, but in the case of the transistor it is not necessary to supply heat to the electrode corresponding to the cathode in order to produce a supply of electrons whose motion in the crystal can subsequently be regulated by the control electrode.

The transistor possesses a number of distinct advantages when compared with a vacuum triode, from an engineering point of view, and we may enumerate them as follows:

1. Transistors may be operated at much lower power levels than vacuum triodes. In the first place there is no requirement for cathode heater power; secondly, it turns out that the bias voltages required are very much lower for a transistor than for a triode. A further incidental advantage in some applications arises from the finite time required for the cathode of a vacuum tube to reach operating temperature after switching the equipment on; there is no such delay in transistorized equipment.

2. Vacuum tubes (valves) are inherently less reliable than transistors, principally on account of the heated cathode in the former. Whenever solids are heated to temperatures of the order of 700°C or above it becomes possible for atoms to move around to some extent, by a process of solid-state diffusion, and eventually to lower the thermionic emission efficiency of the cathode. There are no such effects in transistors, which are generally operated within a range of temperature up to 150°C above room temperature, depending on the amount of powe being dissipated under amplifying or switching conditions.

3. Transistors have considerable advantages of size and weight over vacuum tubes. It is difficult to manufacture a vacuum tube and enclose it in a cylindrical vacuum envelope which is much under 1 cm diameter, and 2 cms. long. Transistors can be fabricated on a silicon crystal as small as 350 microns (μ) square and 100μ thick, on the other hand.

4. Since they are fabricated from a solid crystal, transistors have greater resistance to mechanical shock and vibratioı than valves, for in the latter there are at least three metal electrodes which must be constrained in position in a vacuum enclosure.

It is a natural consequence of these many advantages that transistors should be used to a greater and greater extent in the design of electronic equipment. There was a gradual transition in this direction in the fifteen years following the invention of the transistor, announced in 1948. The trend to transistorized equipment was not complete, however, because not all the capabilities of vacuum tubes can be emulated by transistors; this is particularly the case in respect of high power output, and

amplification at high frequencies. Vacuum tubes are likely to remain superior in these areas for some time to come. Furthermore, economic considerations have sometimes precluded the replacement of valves by transistors in many applications. For these reasons, and also because it is necessary to maintain a supply of vacuum tubes to replace those failing from time to time in equipment already in use, there are today (1969) approximately equal numbers of transistors and vacuum tubes manufactured each year.

One important consequence of the greater reliability of transistors has been to enable engineers to consider the design of more and more complex equipment, incorporating an ever increasing number of active devices. Evidence of this trend is presented in Fig. 5.2. The way in which this consequence comes about is not too hard to see. Suppose that vacuum tubes have unit probability of failing in 5000 hours of operation. A piece of equipment which incorporates 5000 vacuum tubes will then be expected to cease operation, due to failure of one of the vacuum tubes, at intervals of approximately one hour, on the average. If we wish to design a complex computer incorporating 50,000 vacuum tubes we see immediately that the likely period of operation between failures is only six minutes! Since such a level of reliability is totally unacceptable, we can say that the complex computers of today owe their existence for the most part to the invention of the transistor, with its greatly enhanced reliability. As will be seen in the next Section, an even greater improvement in reliability can be achieved with the use of integrated circuits, by means of which additional modes of failure can be circumvented.

Reference was made in the previous Section to the possibility of devising completely new systems as a consequence of the invention of a new device. One of the interesting properties of a silicon p-n junction, to be described in detail in Chapter 7, is its ability to convert incident light into electrical power with a conversion efficiency of approximately 10 per cent. Solar energy, in the form of sunlight,arrives at the earth's surface at a rate of approximately 1 kW per square metre normal to the direction of the sun. A one-square-metre array of silicon "solar batteries" is therefore capable of providing useful electrical power amounting to 100W when sunlight falls on it.

Present satellite communication systems would be quite impracticable but for the power available from solar batteries. Such systems were first contemplated early in this decade, and initially a large aluminized sphere, Echo I, was used to reflect signals from a transmitter on earth and thus to provide intercontinental line-of-sight communication at microwave frequencies. Severe technical difficulties were encountered, arising in part from the low level of the signal power reflected from the satellite. Today's commercial systems are able to receive, amplify and re-transmit on board the satellite, the necessary electrical power being supplied by solar batteries. The development of transistors, with their low power requirements, and of solar batteries, has therefore led to a completely new system.

The reader may well enquire why it is that solar batteries, using "fuel" of zero cost, have not replaced the more conventional methods of electrical power generation on the earth's surface itself. The reasons why this replacement has not taken place are, of course, entirely economic. Silicon solar batteries are at present rather expensive to manufacture, but even if the initial cost of solar batteries were substantially reduced one is still faced with the fact that electrical power is by no means consumed only when the sun is shining. An essential element of such a system will therefore be a cheap means of storing electrical energy. If such a means is devised in the future, this may well be the genesis of another completely new system of power generation.

5.4 The revolution in electronics

We have seen that the invention of the transistor, and its subsequent development, has produced far-reaching changes in the design of electronic equipment. Because of the increased reliability of the active devices, engineers are able to contemplate the satisfactory construction of equipment of a much higher degree of complexity than could be undertaken with the use of vacuum tubes. The lower consumption of electrical power by transistors has similar consequences also. But changes in design concepts of even greater magnitude than these have arisen in the past five years, with the development of silicon integrated circuits. It is with the revolution constituted by these changes that we shall be concerned in this Section.

In the previous Section it was mentioned that a silicon transistor could be fabricated in, or on, the surface of a silicon crystal of square cross-section, 350μ on edge. This dimension was actually specified for the reason that it is the minimum size of crystal that can conveniently be handled in manufacturing processes. (Such processes are naturally carried out under a microscope, in view of the small dimensions of the crystal). As will be shown in Chapter 8, it is actually possible to fabricate many transistors on a crystal of silicon of these dimensions, and to interconnect the transistors subsequently by an appropriate pattern of metal placed in the form of a thin film on the surface of the crystal. Since we may also at the same time fabricate resistors and capacitors as an integral part of the small silicon crystal, we have available a technique for manufacturing a complete integrated electronic circuit, containing up to 20 components, on a crystal whose dimensions are the smallest that can be handled with ease.

These developments have indeed proved to be revolutionary. As we should expect, the possibility of fabricating entire circuits, thousands at a time on a single slice of silicon crystal, leads to substantial economies in cost provided that there is a market to absorb the particular circuits produced. It also turns out to be the case that integrated circuits are inherently more reliable than circuits constructed of discrete components which are soldered together in the conventional fashion. Integrated circuits thus enable engineers to undertake the design, with satisfactory reliability, of even more complex equipments than were possible with transistors and other discrete components.

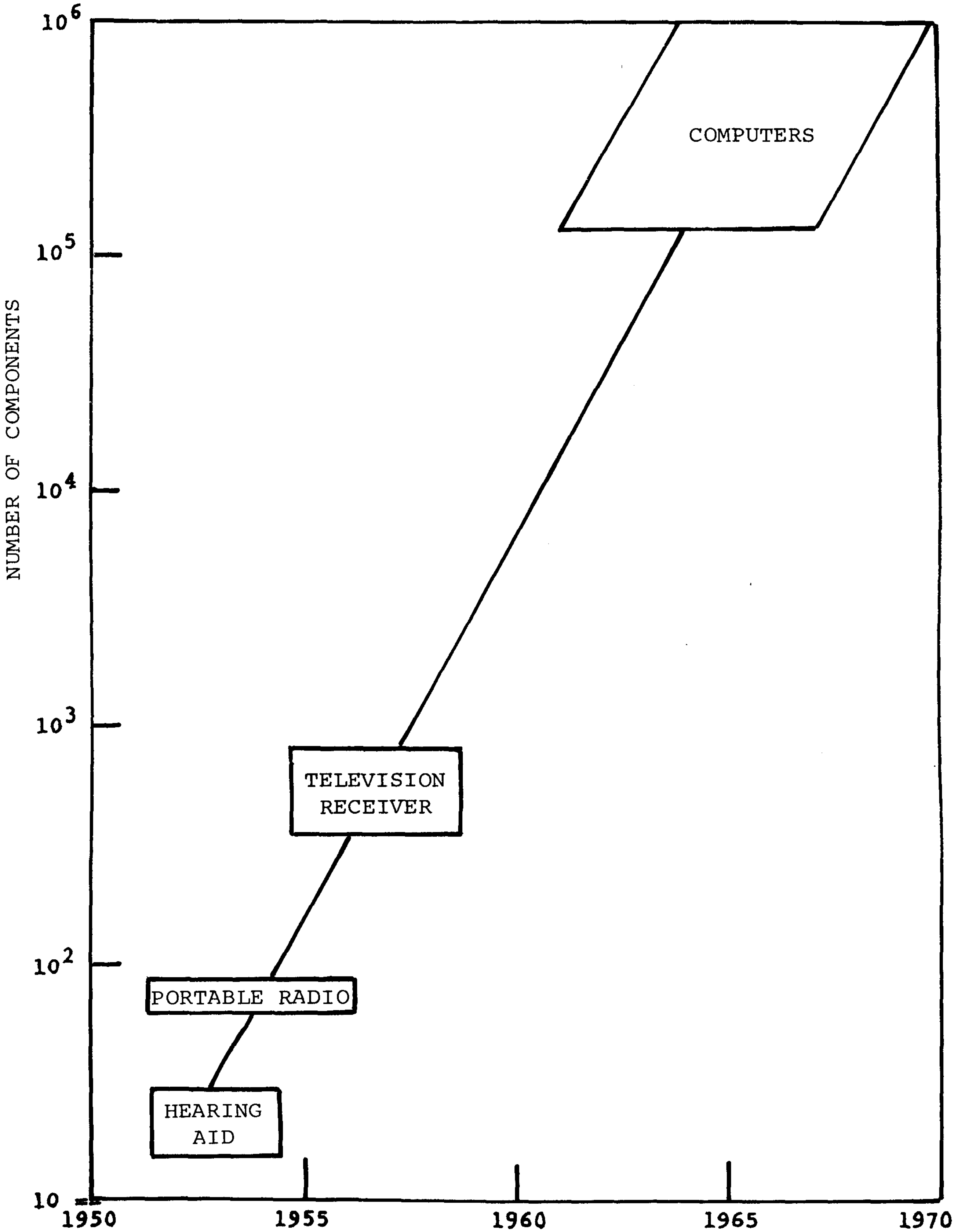


FIGURE 5.2 : TREND WITH TIME OF THE COMPLEXITY OF TRANSISTORIZED EQUIPMENT

From the individual design engineer's point of view, perhaps the most exciting and interesting aspects, however, are the creative opportunities opened up by integrated circuits. No longer is the design of a circuit necessarily a process of minimising the number of transistors to be used; one is instead faced with an available area of silicon crystal on which resistors, capacitors, diodes and transistors can be fabricated to a minimum total of twenty or so, and in any chosen ratio. Furthermore, because it has not proved to be possible to devise satisfactory inductors as integral elements of a silicon crystal, new ways of carrying out particular circuit operations (e.g. frequency-selective amplification) have had to be devised. This is indeed a revolution in electronics, one which is still in progress and to which graduating electrical engineers can look forward to contributing.

5.5 The need for a training in fundamentals

From what has been written to this point in Chapten 5, important general conclusions can be drawn about the profession of electrical engineering.

In the first place, what has happened to electronics since the invention of the transistor in 1948 points up the rapidity with which new inventions and processes are incorporated in modem manufacture. That period, for example, has seen the rise and fall firstly of the point-contact transistor, and then of the germanium junction transistor. There have been other exciting developments as well in the electronics industry, also marked by extreme rapidity of the transition from laboratory to factory.

Another conclusion that may be drawn, when one reflects, for example, on the completely different physical phenomena taking place in a vacuum tube and in a transistor, is the versatility in basic understanding which is likely to be required by a recent graduate entering the profession of electrical engineering. It is probable that undergraduates who read this book will, in their subsequent careers, be faced with the problem of utilizing devices, or making use of techniques, which have not yet been invented. It is for this reason that undergraduates are urged to achieve as wide an understanding as possible of fundamental physical phenomena. This, together with engineering intuition and imagination, will surely enable him to make an effective contribution.

5.6 Problems

1. (i) How many squares 350μ on edge may be obtained from a circular slice cut from a silicon crystal of 5 cm diameter?

   (ii) If it is possible to fabricate 20 components on each of the above squares, how many components can be fabricated simultaneously on the above silicon crystal?

   (iii) How does this number compare with the numbers of components depicted in Fig. 5.2?

   $$(1\mu = 10^{-6}\text{ m})$$

2. (i) What is the total solar power incident at noon on the roof of a house situated at the equator, whose area is 1000 ft ^{2}?

   (ii) If the roof of the house were completely covered with a solar battery of efficiency only 1%, what electrical power would be available to the householder?

5.7 Suggested Further Reading

W. C. Hittinger and M. Sparks: "Microelectronics", Scientific American *213*, No. 5, p.56, November, 1965.

Chapter 6

MATERIALS

.....We are such stuff
As dreams are made on, and our little life
Is rounded with a sleep.

Shakespeare, The Tempest.

Chapter 6

MATERIALS

6.1 Introduction

In Fig. 6.1 is shown the conventional division of matter into the three states: solid, liquid and gas. Because our principal interest here is in electrical properties, we shall focus our attention on the solids. Nevertheless it should be noted that electrical engineers do indeed make use of materials in the form both of liquids and gases; for example, liquid insulators may be used in the manufacture of transformers to increase the rate of heat removal (by convection) from copper conductors, while gases may be employed as insulators.

A fourth "state" of matter – plasma – is also indicated in Fig. 6.1. The term was first used to describe a highly ionized gas, such as is found in fluorescent gas discharge lamps, whose properties differ markedly from those of a non-ionized gas because of large equal densities of positive and negative charge-carriers. The many interesting properties of gas plasmas find wide application in electrical engineering, as also to some extent do the properties of plasmas of electrons and holes found in semi-conductors. Unfortunately space precludes a detailed discussion of plasmas.

If we make a microscopic examination of common salt (sodium chloride, NaCl) we find that the crystalline particles of the material tend to have similar cubic shapes, even though they may differ in size. If salt is dissolved in water and allowed to recrystallize, we again find small cubes, with flat faces. Even when crystals are broken up into fragments, these also tend to have the same cubic shape. From all these observations it can be concluded that in sodium chloride the atoms of sodium and chlorine have a regular, periodic arrangement in space. This conclusion has been confirmed by observations of the diffraction of X-radiation by the regularly arranged planes of atoms in crystals of NaCl, and of many other materials. The X-ray observations give a value of 2.8 Å (1 Å = 10^{-10} m) for the interatomic spacing in NaCl.

Not all solids are crystalline, however. Common glass, for example, which is a mixture of oxides of silicon, sodium and other elements, solidifies in a form in which there is no long-range regularity in its structure. We call such materials amorphous. It is possible for some materials to exist in either form: for example crystals of quartz (α - SiO_2) when melted and re-solidified are found to have been transformed to the amorphous form known as (fused) silica; the chemical composition SiO_2 is retained, but the long-range order of the crystal has been destroyed.

6.2 Classification of crystals

It is convenient for many purposes to classify crystals according to properties of particular interest. Crystallographers, for example, are primarily interested in the symmetry relations of the arrangement in space of the atoms or molecules making up the crystal, and classify accordingly. From the point of view of the electrical engineer, interested as he is primarily in the conductivity and other electrical properties of crystals, it is convenient to classify crystals according to the forces which bind the constituent atoms together. This brings us rather naturally to a consideration of the distribution throughout the crystal of electrons, for electrostatic forces are the ones principally concerned in forming bonds in a crystal, and in order that a current may flow in a crystal there must be some re-arrangement in the crystal of charge-carriers, i.e. of the electrons.

Let us proceed then to enumerate the various classes of bonding mechanism that have been observed in crystals.

6.2.1 Ionic crystals.

This is perhaps the simplest of the crystal binding mechanisms to visualize. Consider NaCl as a typical ionic crystal. The sodium atoms each have a single valence electron in the outermost electron "shell" of the atom, while chlorine atoms have an outer shell which is almost complete, lacking one electron. In solution in water, for example, there is a tendency for the sodium atoms to lose their valence electrons and become positively charged ions (Na^+) with a closed-shell structure, and for the chlorine atoms to take up an electron and become negatively charged ions (Cl^-), also with a closed-shell structure, i.e. with a charge distribution which is spherically symmetrical.

On crystallization, there are forces of attraction between oppositely charged ions, and of repulsion between ions carrying like charges. The solid crystal NaCl forms with the structure shown in perspective in Fig. 6.2. There is a nett binding force holding the crystal together, because the Cl^- ion at the centre of the structure shown is subject ot attractive forces between it and its 6 nearest-neighbour Na^+ ions, which outweigh the forces of repulsion between the central Cl^- ion and its12 next –nearest neighbour ions of like charge. We recall that the electrostatic forces of attraction and repulsion are inversely proportional to the square of the separation between ions.

A feature of ionic crystals such as NaCl is that there is a very low probability of finding electrons in the space between the ions. Consequently ionic crystals are insulators, for there are no electrons available to transport charge through the crystal.

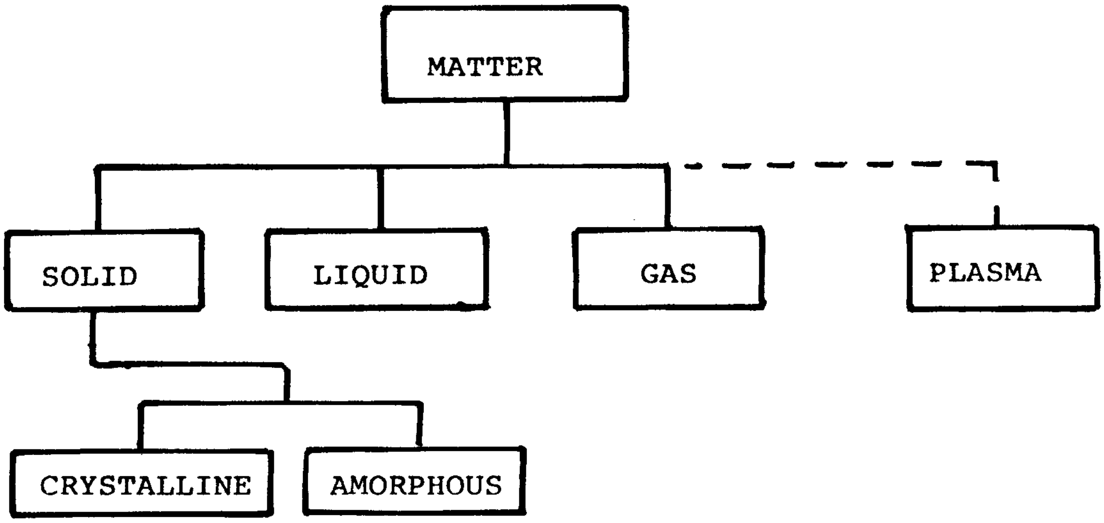


FIGURE 6.1 : THE STATES OF MATTER

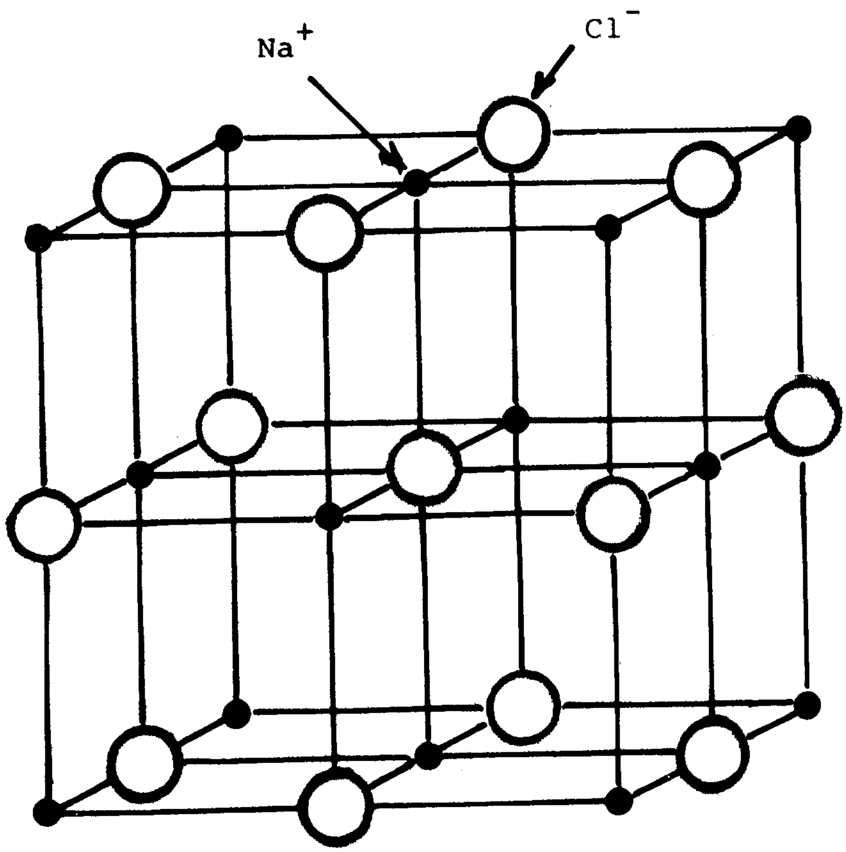


Figure 6.2 : STRUCTURE OF A SODIUM CHLORIDE CRYSTAL

6.2.2 Metallic crystals

Atoms of elements which form metals, of which sodium is a simple example, have one or more valence electrons which are readily detached from each atom as it moves into a growing metallic crystal on solidification. The resultant picture is one of positive ions, regularly arranged in space and "floating in a sea of electrons", as it has been picturesquely described. Once again the forces of repulsion between ions of like sign are outweighed by the attractive forces between ions and the near-uniform distribution of electrons between the ions, and there is a nett force of cohesion as a result. There is a number of possible arrangements in space of the ions, as exemplified by metals such as copper, zinc and chromium, to name three; a schematic diagram of a metal structure is shown in Fig. 6.3, with the distributed electrons shown in colour.

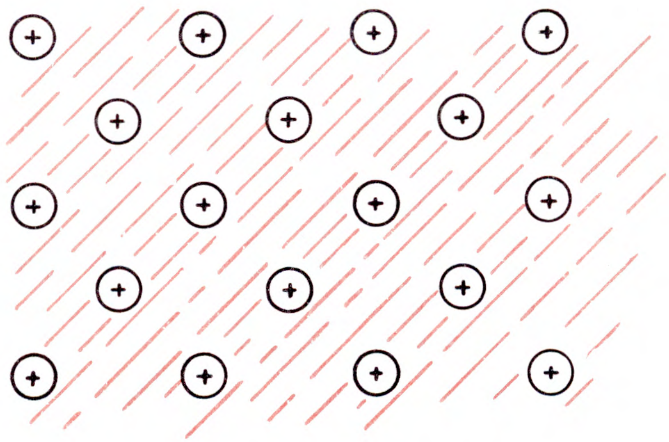
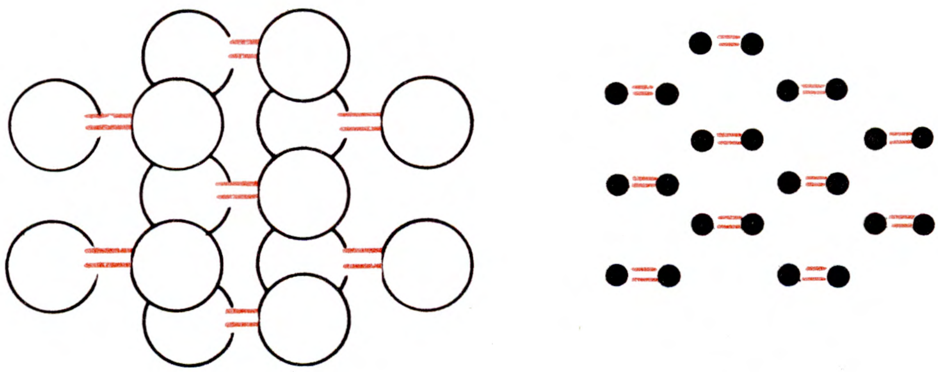


FIGURE 6.3 : STRUCTURE OF A METAL (SCHEMATIC).

It is clear from Fig. 6.3 that the electrons are no longer associated with particular atoms in a metal. There is a high probability that an electron will be found in the space between the ions of the crystal, and this contrasts strongly with the situation in an ionic crystal. This contrast is directly related to the observed contrast in the electrical conductivities of metals and ionic crystals: the distribution of electrons throughout a metal, ready to transport electric current in the presence of an electric field, gives rise to conductivities of approximately 10^5 ohm $^{-1}$ cm^{-1} associated with insulating ionic crystals. Metals are good conductors at all temperatures.

6.2.3 Molecular crystals

It is known that hydrogen gas consists of molecules H_2, each molecule consisting of two protons (the hydrogen nuclei) covalently bound together by two electrons. As the temperature of hydrogen is lowered at atmospheric pressure it first liquefies at 20°K, and then solidifies as a crystal at 14°K. In the crystal structure, shown in Fig. 6.4 with the electron valence bonds in colour, it can be seen that the molecular structure has been retained in the solid. The binding forces are weak electrostatic forces between neighbouring molecules. Their weakness is indicated by the relatively small amounts of thermal energy required to melt the crystal at only 14°K.



PERSPECTIVE VIEW SCHEMATIC

FIGURE 6.4 : THE STRUCTURE OF SOLID HYDROGEN.

From the schematic diagram of Fig. 6.4 it is clear that the electrons in crystalline hydrogen, a molecular crystal, are essentially located within each molecule and are unable to move through the crystal under the influence of an externally applied electric field. Molecular crystals therefore tend to be electrical insulators.

6.2.4 Valence crystals

In the fourth column of the Periodic Table we find, in order, the elements C, Si, Ge, Sn, Pb. Each of these atoms has four valence electrons surrounding a closed electron shell. When atoms of tin and lead are brought together they form a metal, but germanium and silicon (and carbon, in one of its modifications - diamond) solidify from the melt with the crystal structure shown in Fig. 6.5(a). In this valence crystal, each ion is covalently bound to its four nearest neighbours, the bonds consisting of two valence electrons, one from each atom. A valence crystal is essentially a giant molecule, with the valence electrons (shown in colour in Fig. 6.5) located in space in a regular array.

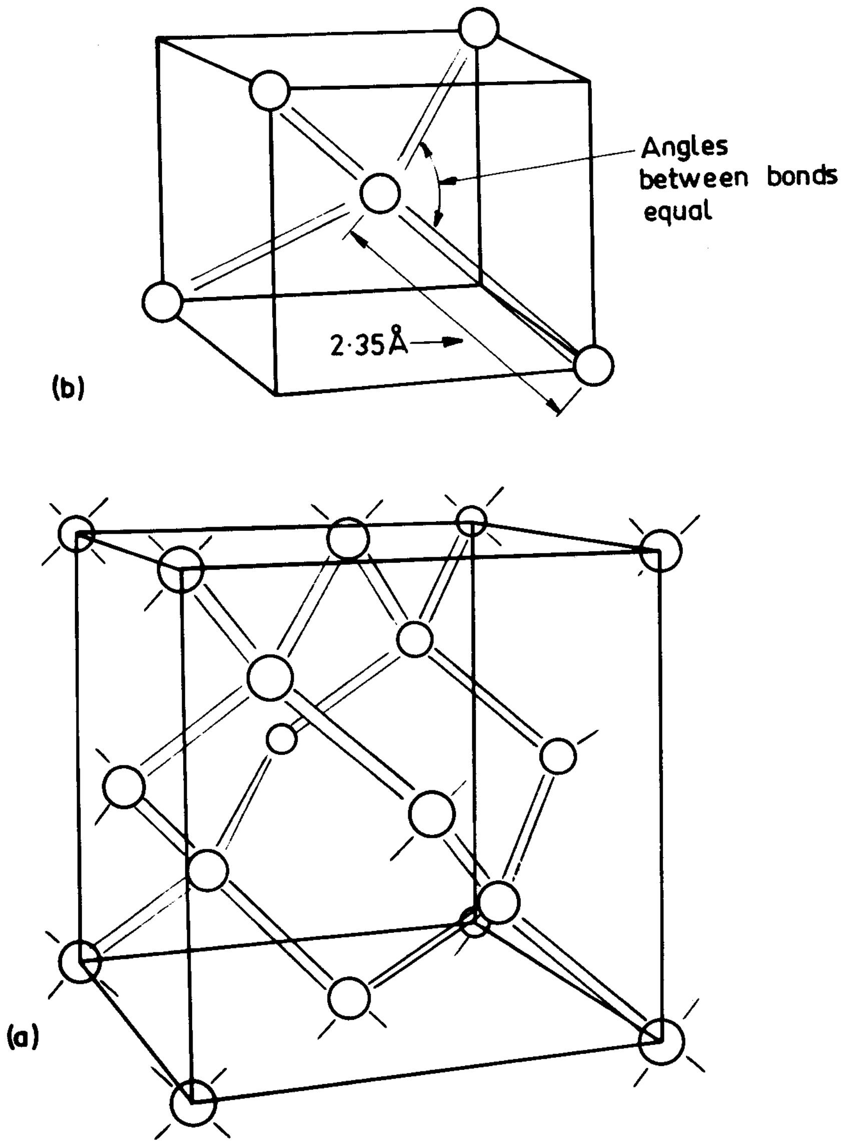


FIGURE 6.5 : (a) COVALENT CRYSTAL STRUCTURE;
(b) NEAREST-NEIGHBOUR RELATIONS IN SILICON.

The structure of a valence crystal differs from the molecular crystal of Fig. 6.4, where the valence electrons play only a small role in holding the molecules together. The binding forces in a silicon crystal are strong, as testified by the high melting point (1414^{o}C) compared with that of hydrogen (14^{o}K).

In order to assist the reader in visualizing the spatial nearest-neighbour relationships in the silicon crystal of Fig. 6.5(a), an element of that structure is shown in detail in Fig. 6.5(b). There is one ion located at the centre of a cube, and its four nearest neighbours are located equally distant at the cube corners shown. It may readily be shown that the four valence bonds to the central ion make equal angles with one another.

The electrical conductivity of valence crystals, and of silicon, is discussed in Section 6.3.2.

6.3 Material properties

In the design of electrical and electronic equipment, engineers naturally make use of virtually all the known properties of materials. In this introductory description of materials it is not possible to elaborate in great detail. We focus our attention primarily on electrical conductivity, for that leads naturally to the important properties of semiconductors,and of semiconductor devices, which have already been shown to have a strong influence on the directions in which electronic engineering is currently making progress. Other properties will subsequently be touched on briefly in Section 6.3.3.

6.3.1 Electrical conductivity

Consider a metal, in which we have electrons free to move throughout the periodic crystal lattice. The electron density will be of the same order as the number density of atoms in the crystal, say $10^{23}\,cm^{-3}$.

If we apply a potential difference across portion of the metal crystal, a current will flow. The electrons are accelerated by the electric field E associated with the potential difference, and in a direction of course opposite to that in which the current is said to flow, according to a convention of long usage. The velocity of each electron does not increase indefinitely when the field E is applied: the electrons are randomly scattered by departures of the crystal lattice from perfect periodicity. Such imperfections may take the form of impurities, or of defects in the crystal structure, but at room temperature the departures from perfection are principally the thermal vibrations of the ions of the crystal about their mean positions in the lattice.

The force on each electron may be written:

$$(-q)\,E = m_n\,a,$$

where $-q$ is the charge on an electron of mass m_n, and a is the acceleration of the electron due to the electric field E. Let us make the simplifying assumption that each electron has a time of flight τ between collisions with the lattice, and that after each collision the electron starts with zero velocity. Then the distance travelled by each electron between collisions is $\frac{1}{2} a\tau^2$, and the average velocity due to the presence of the electric field is therefore

$$\bar{v}_n = \text{distance/time} = \tfrac{1}{2} a\tau^2/\tau = \tfrac{1}{2} a\tau.$$

Substituting the value of a given above,

$$\bar{v}_n = -\,(q\tau/2m_n)\,E.$$

We have thus shown that the average velocity, or drift velocity of the electrons, $\bar{v}_n$, is proportional to the electric field strength E; we write

$$\bar{v}_n = -\,\mu_n E,$$

where the constant μ_n is known as the electron mobility.

The density of current carried by the electrons is given by the flux of electrons (number per unit area per unit time) multiplied by the charge on each electron, that is

$$i = n\bar{v}_n\,(-q),$$

whence $i = nq\mu_n E$;

or $i = \sigma E$

where the conductivity $\sigma = nq\mu_n$.

On the basis of the simplifying assumptions made above we have thus derived Ohm's Law, and have obtained a theoretical expression for the electrical conductivity of a metal. The true situation is complicated by the random thermal motion of the electrons, but we shall not discuss such complexities further. The conductivity is seen to be dependent both on the density of charge-carriers and on their mobility.

6.3.2 Conduction in valence crystals

In the previous Section 6.2 it was concluded that ionic and molecular crystals are insulators, while metals are good conductors at all temperatures. What of the conductivity of a valence crystal?

According to the model of a silicon crystal shown in Fig. 6.5, the application of an electric field would not lead to conduction as the electrons, located in valence bonds throughout the crystal, are not free to move. It is important to note, however, that the model of Fig. 6.5 depicts the situation at the absolute zero of temperature, for the atoms are not vibrating about their mean positions. As the temperature is increased the thermal vibrations of the atoms will in fact disrupt a small proportion of the valence bonds; electrons are then set free to move through the crystal and to contribute to electric conduction. The density of electrons, and hence the conductivity, increases strongly with increasing temperature.

For convenience of discussion the valence crystal structure of Fig. 6.5 may be shown schematically in two dimensions (Fig. 6.6). Each atom has four nearest neighbours to which it is covalently bound. The disruption of a bond to create an electron free to move through the crystal is also illustrated.

It can be seen immediately from Fig. 6.6 that a further mechanism exists for the transport of current through a valence crystal, over and above the conduction due to the electrons freed from the valence bonds. The electric field also acts on the electrons still remaining in the valence bonds, and it is possible for an electron to move from an adjacent bond into the vacant bond position. By a sequence of such movements, it is seen that electrons in the valence bonds can also contribute to conduction when there is a vacancy present. Alternatively, attention may be fixed on the vacancy itself; the electric field causes the vacancy to move, and this movement is in a direction opposite to that of the electron freed from the valence bond, that is, in the direction a positive charge would move.

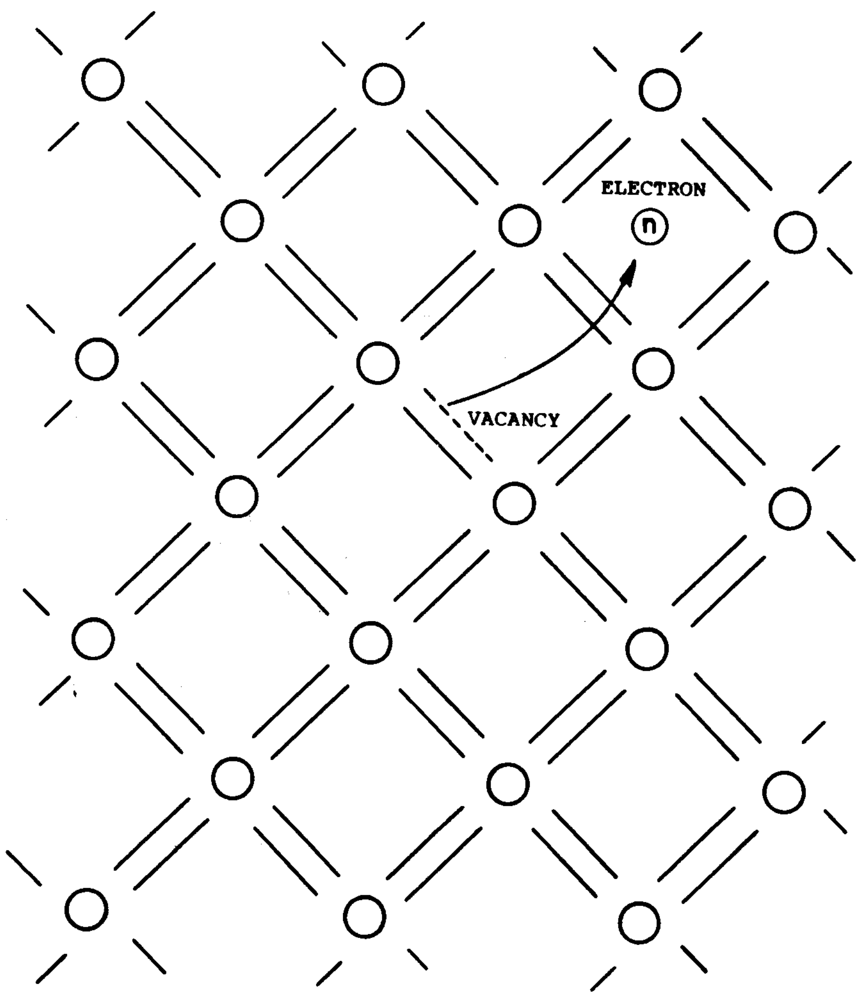


FIGURE 6.6 : TWO-DIMENSIONAL REPRESENTATION OF A VALENCE CRYSTAL.

Thus there are two essentially independent modes of conduction in a valence crystal: conduction by the electrons freed from the valence bonds (density n), and by the vacancies in the valence bond structure which we call "holes" (density p). The concept of a hole is simply a convenient method of describing the motion in an electric field of all the valence electrons in the crystal when there is one such electron missing; the density of holes is denoted by p to remind us that the holes move in an electric field as though they were positively charged.

The above discussion is a simplification of a much more elegant and informative treatment using wave mechanics. It does however give us some insight into the processes of conduction in valence crystals such as silicon and germanium, in spite of the imperfections of such a simplified treatment.

If the electron and hole mobilities are respectively μ_n, μ_p, the conductivity of a valence crystal is given by

$$\sigma = nq\mu_n + pq\mu_p.$$

In the case of the pure or intrinsic crystal of Figs. 6.5 and 6.6 the electron and hole densities are equal, since they are created in pairs. In the next Chapter we see that it is possible to vary the electron and hole densities by incorporating impurities in a valence crystal.

The electron and hole density

$$n = p$$

in an intrinsic valence crystal is determined by two competing processes at any particular temperature. On the one hand we have the continuous creation of electron-hole pairs by the thermal vibrations of the lattice, while on the other it must be recognized that electrons and holes may spontaneously recombine, the energy involved being transformed into heat or radiation. It turns out that the equilibrium densities n and p increase exponentially with temperature, the mobilities being relatively weakly dependent on temperature. The densities n, p are very much less than the number density of atoms in the crystal, so that the conductivity σ, as well as being an exponential function of temperature, is intermediate between that of metals and that of insulators. This is the observed pattern of behaviour in the materials known as semiconductors.

6.3.3 Engineering properties

Regrettably it is not possible in this book to do more than follow up in detail the electrical properties of materials, and to show how these properties have been applied to the invention of useful devices, particularly in the semiconductor field. Nevertheless the story would be quite incomplete if we did not refer in passing to other aspects of material properties.

We have seen that solid crystals are either insulating, conducting or semiconducting, to an extent dependent on the binding mechanism. In the case of some metals and metal alloys there is an abrupt transition to infinite conductivity as the temperature falls below a transition temperature, usually below 20°K. A current set up in a closed loop of a superconductor, as such materials are known, will flow indefinitely, provided the temperature is not allowed to rise above the critical temperature, and provided also that the magnetic field strength where the current is circulating is kept below a critical value. There have been recent engineering applications of superconductors to the construction of high-field magnets, which are of some importance, and further engineering advances can be expected as materials of higher transition temperature are formulated in the future.

In all materials other than superconductors, dissipation of energy is associated with the flow of current. In order to dispose of heat generated in this way, provision must be made for its removal by thermal conduction. The thermal conductivity of materials is therefore of great importance also. Since thermal conductivities are finite, increases in temperature can be expected in any devices (other than superconductors) in which current flows and thus engineers have concern also for the temperature-dependence of properties; e.g. the dependence with temperature of mechanical strength, ferromagnetism, cathodoluminescence, or whatever is of importance to the application concerned. Ferromagnetic materials, to take one example, lose their spontaneous magnetic moment and become paramagnetic above a temperature characteristic of each material (known as the Curie temperature), and care must be taken therefore to ensure adequate dissipation of heat to prevent the Curie temperature being reached.

In materials in which non-uniformities of compositon have been introduced by design, such as the cathode of a vacuum tube, or a p-n junction in a semiconductor, another fundamental consequence of increases in temperature must be taken into account. At temperatures above a few hundred degrees Centigrade the thermal vibrations of atoms become so intense that some of them leave their positions in the crystal lattice, creating vacancies, and diffusing through the crystal under the influence of any gradients in concentration which may be present. The desired non-uniformities in composition thus tend to become less pronounced, and this is of course the mechanism of failure in cathodes of vacuum tubes which was referred to in Chapter 5.

Solid-state diffusion processes of this kind may also be employed to advantage, however. Heating a semiconductor crystal in the presence of impurity elements leads to solid-state diffusion of the impurity into the crystal, and one is in this

way able to produce a distribution of impurity in the crystal which is of use in device fabrication, as in the fabrication of silicon integrated circuits (see Chapter 8). The rate of diffusion increases exponentially with temperature, so that with control of time and temperature one is able to achieve an extraordinarily close control of the distribution of impurity in the crystal.

6.4 Problems

1. Given that the atom spacing in a silicon crystal is 2.35 Å, with the aid of Fig. 6.5 show that the number density of silicon atoms is $5.10^{22}cm^{-3}$.

2. By taking plane sections of the cube of Fig. 6.5(b), prove that the angle between any two valence bonds in a silicon crystal is 109.5^o.

3. Construct a model of portion of a silicon crystal, using lozenges to represent the atomic ions and toothpicks to represent the valence bonds. Make use of Fig. 6.5(b) to achieve the correct bond angles.

4. Calculate the electrical conductivity of an intrinsic silicon crystal at 300^oK, when it is known that the electron and hole densities are $1.5 \times 10^{10}cm^{-3}$, and the mobilities are respectively 1350 and 480 $cm^2V^{-1}s^{-1}$. (The magnitude of the charge on an electron is $1.6 \times 10^{-19}C$).

6.5 Suggested Reading

"Materials" – A Scientific American Book.

(W. H. Freeman and Co., San Francisco, 1968)

(Reprint of *The Scientific American, 217,* No. 3, September, 1967)

CHAPTER 7 – SEMICONDUCTORS

"Behold, the half was not told me."
1 Kings, 10:7

CHAPTER 7 – SEMICONDUCTORS

In Chapter 5 the semiconductor revolution was described in somewhat general terms. It is now proposed to discuss semiconductor devices, so that a deeper understanding may be achieved of the overwhelming influence, in bringing about this revolution, of the semiconducting elements germanium and silicon. We first investigate the electrical properties of semiconductors, placing particular emphasis on the properties of slightly impure semiconductors, and on their non-equilibrium properties.

7.1 Intrinsic semiconductors

The conduction mechanisms in pure, or intrinsic, semiconducting valence crystals were described in Section 6.3.2. Briefly recapitulating that discussion, both germanium and silicon crystallize with the structure shown in Fig. 6.5(a): a regular array of electrons in valence bonds extends throughout the space occupied by the crystal. In an intrinsic semiconductor at non-zero temperatures some of the valence bonds are disrupted, giving rise to equal densities of conduction electrons and holes. Electrons and holes are charge-carriers of negative and positive sign respectively, moving in opposite directions in the presence of an electric field to make quite independent contributions to conduction in the crystal. The electrical conductivity is thus given by the sum of two terms:

$$\sigma = nq\mu_n + pq\mu_p .$$

In an intrinsic semiconductor the electron and hole densities n and p are equal, and increase strongly with increasing temperature. We write, for the case of an intrinsic semiconductor,

$$n = p = n_i(T) ,$$

where the subscript i denotes intrinsic. In the case of silicon, for example,

$$n_i(300^\circ K) = 1.5 \times 10^{10} \text{ cm}^{-3}.$$

7.2 Extrinsic semiconductors

The electrical conductivity of any semiconductor may be profoundly modified by the presence in the crystal of quite small amounts of impurity, in the following way. Consider for example a silicon crystal which has grown from a melt in which there was present a small amount of arsenic also. Each arsenic atom has five valence electrons, surrounding a closed electron shell. In the growth of the crystal arsenic atoms are incorporated substitutionally, taking the place of occasional silicon atoms in the lattice, as shown schematically in Fig. 7.1. Four of the five valence electrons are shared with neighbouring silicon atoms to form covalent bonds, and the fifth electron remains bound to the arsenic ion by electrostatic forces, at very low temperatures.

The specific dielectric constant of silicon (ϵ = 11.7) is quite high, and as a consequence the energy required to detach the fifth valence electron from the arsenic ion is small, being only 0.04 eV. At temperatures above $50^\circ K$ virtually all such electrons are detached from their parent arsenic ions, being then free to move through the crystal as conduction electrons. The positively charged arsenic ion remains behind, fixed in position in the lattice. In this way arsenic impurity atoms (or atoms of any of the elements P, Sb, Bi of the 5th column of the Periodic Table) can increase the equilibrium density n of conduction electrons in the crystal. Such elements are referred to as donor impurities, each atom "donating" a conduction electron to the crystal.

The number density of atoms in a silicon crystal is 5×10^{22} cm^{-3}. If only one in every 10^8 of these atoms is replaced by a donor impurity there will be 5.10^{14} donor atoms in each cm^3, and approximately the same density of conduction electrons. Such a density is very much greater than the density of electrons in pure silicon at room temperature, namely $n_i(300^\circ K) = 1.5 \times 10^{10}$ cm^{-3}. Thus the conductivity of silicon is increased some 33,000 times at room temperature by the addition of only 10^{-8} atom fraction of donor impurity. Furthermore, in the arsenic-doped crystal the flow of current is almost entirely due to electrons, since they are so much more numerous; such a crystal is said to be n-type.

We have thus demonstrated a profound influence of Column-V impurities on the electrical conductivity of silicon. Semiconductors influenced in this way are said to be extrinsic, in contrast to intrinsic (pure) semiconductors.

As might be expected, impurity elements with only three valence electrons (column III of the Periodic Table) also lead to an increase in conductivity when incorporated in an otherwise pure semiconductor, but they do so in this case by increasing the density of holes, p. Consider for example a substitutional atom of boron, shown schematically in the silicon crystal of Fig. 7.1. In this case there are only three valence electrons to share with the four neighbouring silicon atoms,

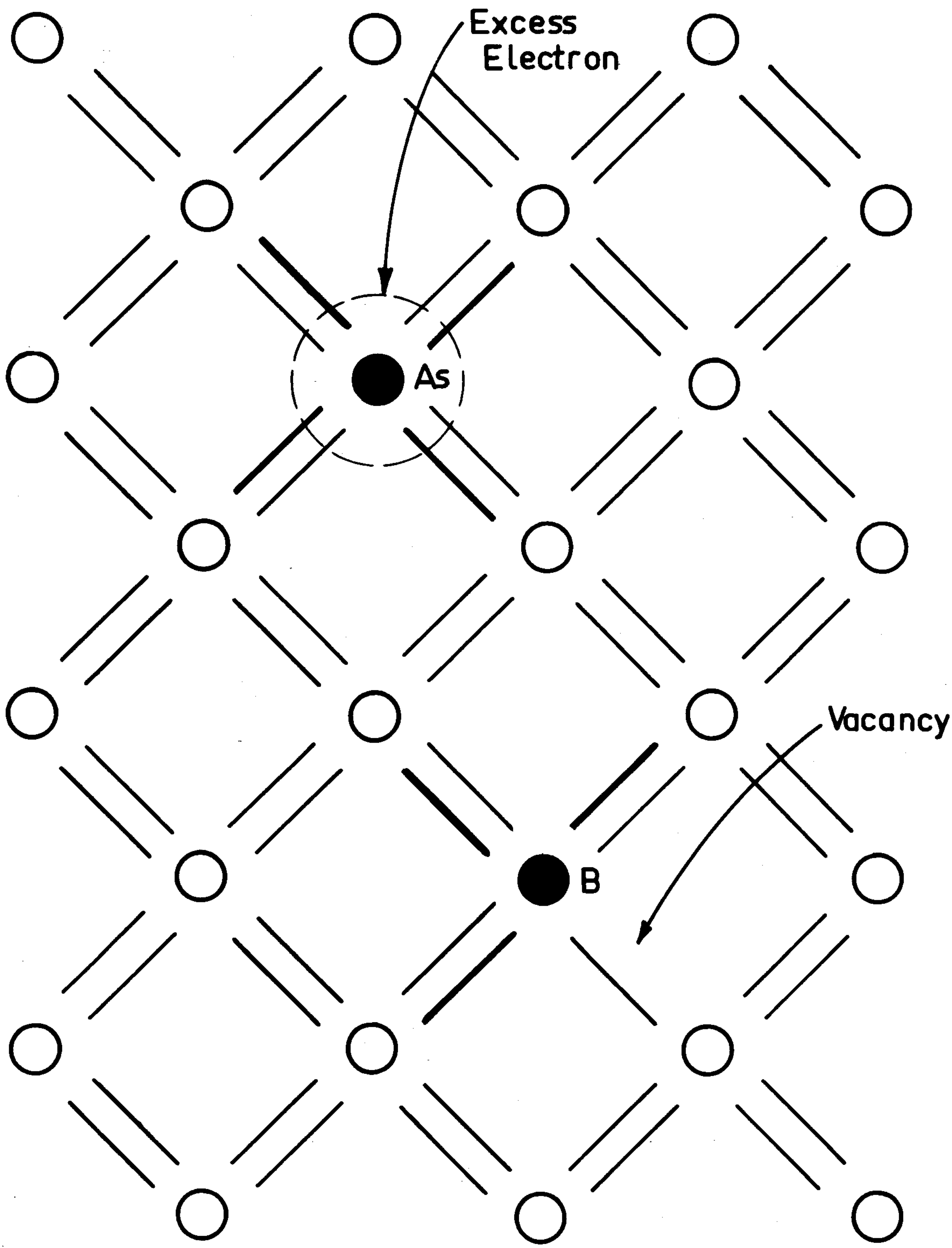


FIGURE 7.1 : ARSENIC AND BORON IMPURITY ATOMS
IN A SILICON CRYSTAL AT LOW TEMPERATURE.

and a vacancy in the valence bond structure is associated with the boron ion. Once again the binding energy is small, and the vacancy is readily detached from the boron ion to become a mobile vacancy, or hole. There remains behind a boron ion plus an additional electron which it has "accepted" to complete its surrounding valence bonds, so that the ionized acceptor impurity (as it is called) carries a negative charge. Semiconductors which contain such acceptor impurities (B, Al, Ga, In or Tl) are said to be p-type, since their electrical conductivity arises almost entirely from the movement of holes (positive charge carriers).

We have seen that very small amounts of impurity from columns III and V of the Periodic Table can strongly influence the conductivity of silicon. In view of the difficulty in purifying silicon to levels less than 10^{10} cm^{-3}, i.e. to less than one part in 5.10^{12}, it is likely that both donor and acceptor impurities are present together in any given crystal. For this reason, and also for reasons connected with the fabrication of silicon transistors (see Chapter 8), it is instructive to calculate the equilibrium densities of electrons and holes in a crystal containing both donor and acceptor impurities, densities N_D, N_A cm^{-3} respectively. We take the impurities to be fully ionized, i.e. $T > 50^oK$.

There are two basic equations involved in the calculation.

In the first place, we note that the addition of donor or acceptor atoms to a semiconductor crystal does not imply any departure from charge neutrality, since the atoms are neutral. The densities of positive and negative charge-carriers in the crystal may therefore be equated, yielding

$$p + N_D = n + N_A.$$

Secondly, it turns out from considerations of thermodynamic detailed balance in a semiconductor in equilibrium that the product of electron and hole densities is constant, independently of the amount of impurity present. Thus we may write

$$np = \text{constant}, = n_i^2(T),$$

since an intrinsic crystal (where $n = p = n_i$) is one such case.

With the aid of these two equations we may calculate the equilibrium electron and hole densities for any given N_D, N_A and temperature. For example, suppose $N_D > N_A$ in a given silicon crystal; substituting for p in the first equation, we find

$$n^2 - (N_D - N_A)n - n_i^2 = 0,$$

a quadratic in n with only one positive solution:

$$n = \tfrac{1}{2}(N_D - N_A)\left(1 + \left\{1 + 4n_i^2/(N_D - N_A)^2\right\}^{\frac{1}{2}}\right).$$

It can be seen that when $(N_D - N_A) \gg n_i$ the electron density is very closely equal to the effective donor density, $(N_D - N_A)$; part of the actual donor density N_D goes toward "compensating" the acceptor atoms present. The compensation of electrically active impurities is made use of in transistor fabrication.

7.3 P-N junctions

An interesting situation arises – one of great importance to semiconductor device physics – when we consider the properties of a monocrystalline semiconductor containing adjacent p-type and n-type regions, formed as a result of a non-uniform distribution of donor or acceptor impurities. A p-n junction arising from abrupt discontinuities in donor and acceptor distributions is illustrated in Fig. 7.2.

Measurements of the hole density p in the crystal of Fig. 7.2 would give values $p \cong N_A$ in the p-region, and very small values indeed (approximately n_i^2/N_D) in the n-region. It must be concluded that some influence associated with the p-n junction is preventing the diffusion of holes from p- to n-region under the influence of the gradient in hole density which exists. There is in fact a contact potential difference V_o between p- and n-regions, the n-region being positive with respect to the p-region and thereby constraining almost all the holes to remain in the p-region. The potential barrier, of height V_o, is at the same time of the required polarity to constrain the electrons to the n-region (see Fig. 7.2). In equilibrium there is an exact balance of the hole currents (and also of the electron currents) in the two directions through the junction. The current that arises from the small number of holes able to surmount the potential barrier and diffuse into the n-region is exactly balanced by a current of holes, thermally generated in the n-region, which diffuse to the potential barrier and are thence accelerated to the left and into the p-region. A similar argument applies to the electrons.

The varying electrostatic potential in the vicinity of the junction has associated with it a space charge layer, also shown in Fig. 7.2. The space charge to left and right of the junction is made up of ionized acceptors and donors respectively;

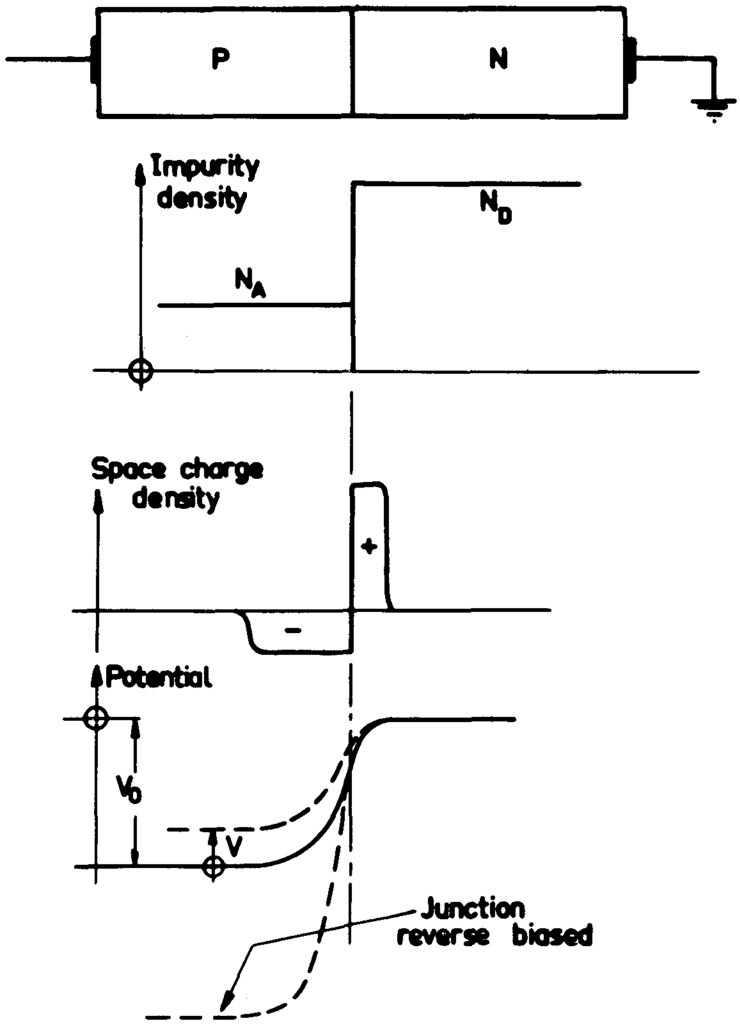


FIGURE 7.2 : SILICON P-N JUNCTION.

the densities of electrons and holes in the space charge layer are negligibly small, and for this reason it is also referred to as a depletion layer. As the depletion layer is a region of low conductivity which supports an electric field, as also is the dielectric in a parallel-plate capacitor of conventional construction, it follows that all p-n junctions in semiconductors have capacitance associated with them.

When metallic contacts are made to p- and n-regions of the junction (as illustrated in Fig. 7.2), and a voltage V applied externally, the balance in the two components of hole and electron current through the junction which exists in equilibrium is destroyed. Suppose a positive voltage V to be applied so as to make the p-region more positive with respect to the n-region, thereby lowering the height of the potential barrier from V_o to $(V_o - V)$. Many more electrons and holes are then able to surmount the barrier and diffuse respectively into the p- and n-regions, where they eventually recombine. The current through the junction therefore increases rapidly (indeed, exponentially) with increasing V of this polarity, and we say that the junction is biased in the forward or easy direction of current flow.

On the other hand, if the junction is biased in the reverse direction by a voltage of reverse polarity (i.e. $V < 0$), the height of the potential barrier is increased. Holes from the p-region, and electrons from the n-region, can no longer surmount the barrier: we are left with the two components of current due to the thermal generation of holes in the n-region, and of electrons in the p-region. In principle, these two components are independent of the applied voltage.

The current-voltage characteristic of a p-n junction is shown in Fig. 7.3, and is given by the equation:

$$I = I_S \left\{ \exp(qV/kT) - 1 \right\} ,$$

where k is Boltzmann's constant; $-I_S$ is the saturation value of I, i.e. the current which flows for negative values of V such

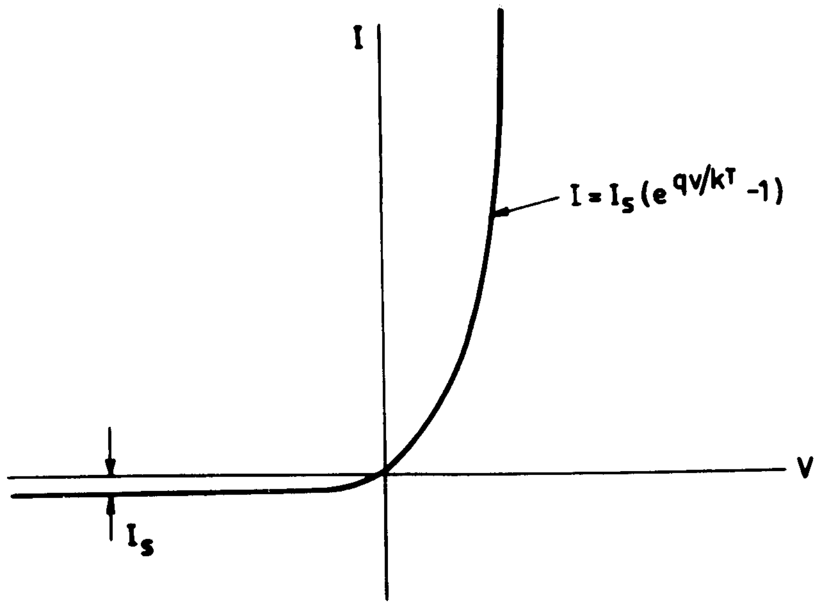


FIGURE 7.3 : CURRENT-VOLTAGE CHARACTERISTIC OF A P-N JUNCTION.

that

$$-V \gg kT/q.$$

(At room temperature, $kT/q = 25$ mV). It is seen that a p-n junction is very nearly an ideal rectifier. We note that the saturation current I_s is a strong (exponential) function of temperature, since it arises from the thermal generation of electron-hole pairs by the vibrating atoms of the lattice.

Light which falls on a semiconductor may also create electron-hole pairs by the disruption of valence bonds, provided the photons have sufficient energy. In silicon the energy required is 1.1 eV or more, corresponding to photons of wavelength less than 1.1μ. If the electrons and holes are created near a reverse biased p-n junction, some of them will be accelerated across the junction by the electric field in the depletion layer and the current will rise above I_s. A p-n junction operated in this mode, so that it "collects" charge-carriers in excess of equilibrium in its vicinity, makes a useful detector of radiation.

Even if no external bias is applied to a p-n junction, the built-in potential barrier is of sufficient height to separate electrons and holes which are created nearby in excess of equilibrium. Holes and electrons pass into the p- and n-region respectively. In this way sunlight incident on the p-n junction of a solar battery creates charge carriers which, when separated by the potential barrier, set up an electromotive force. If a resistive load is connected across an illuminated solar battery a current will flow, and there is a direct transfer of energy from sunlight to electric power dissipated in the load. If the p-n junction is arranged to lie about 1μ below the surface of the silicon crystal, where most of the electron-hole pairs are created by sunlight, conversion efficiencies as high as 10% are readily attained. The importance of solar batteries to satellite communication systems has already been referred to in Chapter 5.

When a negative voltage V is applied to a p-n junction there is a widening of the depletion layer. As the depletion layer widens, the capacitance associated with the junction decreases: the equivalent "parallel plates" are being drawn further apart. Thus a further interesting property of a p-n junction is that it has a capacitance C(V) which is a function of the voltage applied to the junction.

7.4 Junction transistors

As we have seen, the potential barrier in a p-n junction is lowered by the application of a bias voltage in the forward direction. The electrons which are then able to surmount the barrier diffuse into the p-region adjacent to the junction, where

they are in excess of equilibrium. The electrons diffuse further into the p-region, recombining with holes at the same time; we have a steady-state distribution of excess electrons which have been injected electrically into the p-type material.

We have also seen that a p-n junction biased in the reverse direction will collect any excess carriers which arrive in its vicinity. It follows that if a reverse biased junction is placed sufficiently close to a forward biased junction in the same crystal, as shown in the n-p-n structure of Fig. 7.4, the current flowing to the reverse biased junction can be controlled by the current which we allow to flow in the forward biased junction. As will be shown, the n-p-n structure, which we call a transistor, is capable of amplifying a signal.

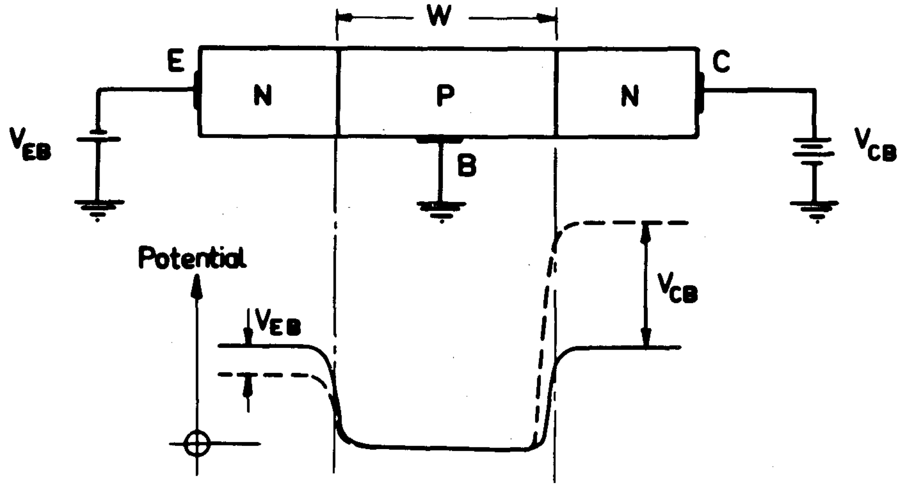


FIGURE 7.4 : NPN JUNCTION TRANSISTOR.

The distribution of electrostatic potential in equilibrium in the n-p-n transistor of Fig. 7.4 is shown by the full line. We put the central base region (p-type)- at zero potential by earthing it. When a positive voltage V_{CB} is applied between the right hand electrode (the collector) and the base, the collector junction is biased in the reverse direction and only the saturation current flows; the increase in collector barrier height is shown by the right hand dashed line. If now a negative voltage V_{EB} is applied between the left hand electrode (the emitter) and the base, the reduced height of the emitter potential barrier allows electrons to flow into the base region, where they are in excess of equilibrium. If the width W of the base region is sufficiently small, i.e. is less than some tens of microns in the case of a silicon transistor, almost all the electrons diffuse to the collector junction before they are able to recombine. We are thus able to enhance the current flowing in the collector-base circuit by means of a current flowing in the emitter-base circuit. In a well designed n-p-n transistor almost all the current flowing through the emitter junction is carried by electrons, almost all of which reach the collector before recombining with holes in the base. The current I_B flowing in the base lead is consequently only one or two percent of the current I_E flowing in the emitter lead.

Junction transistors are often operated in the "common emitter" configuration of Fig. 7.5; a positive voltage V_{BE} between the base and the earthed emitter determines the base current I_B, which in turn controls the much larger collector current. The bias voltage V_{CE} applied between collector and emitter is positive. The distribution of current between the three electrodes is also shown in Fig. 7.5. An input signal current in the base lead causes an amplified current to flow in the collector circuit.

It is clear that there are two possible types of transistor: the n-p-n configuration discussed above, and a p-n-p configuration in which holes are injected into the n-type base by the emitter and diffuse to the collector. In this respect transistors are more versatile than vacuum triodes, in which the current that is controlled can only be carried by electrons.

7.5 Problems

1. Calculate the electron and hole densities in a crystal of silicon at 300°K which contains 10^{15} cm^{-3} substitutional gallium atoms.
(For silicon, $n_i(300^oK) = 1.5 \times 10^{10}$ cm^{-3}).

2. Calculate the ratio of electron and hole densities in a germanium crystal at 300°K containing 5×10^{13} cm^{-3} substitutional indium atoms; take $n_i(300^oK) = 2.3 \times 10^{13}$ cm^{-3} in germanium. Is the crystal n-type or p-type?

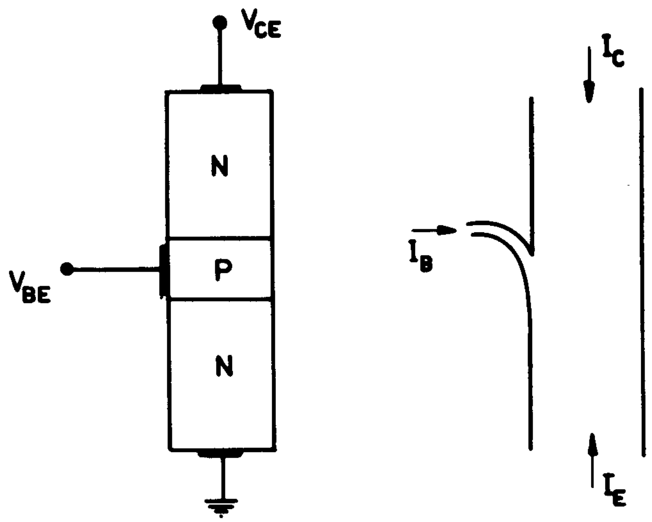


FIGURE 7.5 : NPN TRANSISTOR IN COMMON-EMITTER CONNECTION, AND DISTRIBUTION OF CURRENT BETWEEN ELECTRODES.

3. A crystal of silicon contains 10^{16} cm^{-3} donors and 10^{15} cm^{-3} acceptors; calculate the electron density at room temperature.

7.6 Suggested reading

R.M. Rose, L.A. Shepard and J. Wulff: "The Structure and Properties of Materials – Volume IV: Electronic Properties",

(New York, John Wiley and Sons, Inc: 1966), Chapters 5 and 6.

L.W. Davies and H.R. Wilshire: "Semiconductors and Transistors, and Further Experiments in Electronics".

(Sydney, Amalgamated Wireless Valve Co. Pty. Ltd.: 1965).

CHAPTER 8 – INTEGRATED CIRCUITS

"All are but parts of one stupendous whole
Whose body nature is, and God the soul."
Alexander Pope.

Chapter 8

INTEGRATED CIRCUITS

8.1 Introduction

In the previous Chapter the operation of an n-p-n transistor was discussed in the form of an idealized monocrystalline structure with contiguous n-, p- and n-regions. Nothing has been said to this point on the question of how one might fabricate such a device. Severe technological problems had to be overcome before the first junction transistors were fabricated of germanium in 1951, particularly in view of the close spacing required for the two p-n junctions of the device, and the close tolerances to be met in impurity content of the crystal.

No detailed historical development will be given here of the successively improved forms of junction transistor, culminating in the so-called Planar silicon transistor. Nevertheless it is important to keep in mind that the advanced form of transistor manufactured today has depended for its development on a long series of technological advances in the processing of semiconductor materials, and that these advances in turn have depended on a deepening knowledge of the basic physical properties of materials. Basic physics, process technology and semiconductor electronics have advanced together, and will no doubt continue to do so in the future.

In a sense, integrated circuits were a natural and consequential development of the Planar process for the fabrication of silicon transistors. The advances achieved in switching speed of the new devices, and in reliability, made it essential to consider a more compact and reliable method of interconnection if full advantage were to be taken of the situation. We therefore preface our discussion of integrated circuits with a brief description of the fundamentals of the Planar process for fabricating silicon transistors.

8.2 Solid-state diffusion in silicon

Reference was made in Section 6.3.3 to the diffusion processes which occur in solids at high temperatures. If a crystal of silicon containing a non-uniform distribution of impurity atoms is heated to temperatures above 900° C, the impurity atoms become displaced in the crystal in such a way as to tend to remove any gradient in their concentration. The changes in concentration arise from the random thermal motion of the atoms of silicon and impurity at high temperature.

The same processes occur when a pure silicon crystal is exposed to an atmosphere of impurity atoms at high temperature, for in this case there is a steep gradient in the concentration of impurities at the surface of the crystal. The impurities condense on the surface, and subsequently penetrate the silicon by solid-state diffusion. The ultimate depth of penetration may be very closely controlled by precise control of the temperature of the process, and the time for which it proceeds.

Consider a uniformly doped n-type silicon crystal heated, say, for one hour at 1150°C in the presence of an atmosphere of boron atoms. (We neglect some of the details of the technological process, such as the exact composition of the atmosphere in which the process takes place). At the end of this time, examination would show a distribution of boron atoms (acceptors) at the surface of the crystal given by the curve labelled N_B in Fig. 8.1, which gives the impurity content in the crystal plotted on a logarithmic scale, as a function of distance plotted on a linear scale. It is seen that there is a layer at the surface, 2.5μ deep, in which the density of boron atoms N_B is everywhere greater than the initial donor density N_D, so that we now have a p-type layer on an n-type crystal.

If now a second diffusion step is carried out, but this time in an atmosphere of phosphorous (donor) impurities at a lower temperature, the result is a distribution of donors given by curve N_P in Fig. 8.1. In the new surface layer, of depth approximately 2μ, the donor density is greater than the acceptor density and it is therefore n-type. At the lower temperature, the distribution N_B of boron atoms remains essentially unaltered.

We have thus produced an n-p-n structure in a silicon crystal by two sequential solid-state diffusion processes, which enable close control to be achieved on the width of the p-type layer, and on the impurity densities. Impurity distributions of the type shown in Fig. 8.1 are a little different from the uniform donor and acceptor densities considered in Chapter 7, but there is actually a distinct advantage in having a non-uniform distribution of acceptor density in the base region.

From the equilibrium distribution of electrostatic potential in the n-p-n structure, shown in Fig. 8.1, it can be seen that there is a gradient in potential, and therefore an electric field, in the base region in particular. Consideration will show that this electric field is of the polarity which will accelerate electrons from left to right. In a uniform-base transistor, the electrons travel from emitter to collector by the relatively slow process of diffusion. If we were to make the surface (left-hand) n-region of Fig. 8.1 the emitter of a transistor, however, the electrons would actually be accelerated across the base to the right-hand (collector) region, and their transit times reduced by a factor of 10 or more. Thus the impurity distributions obtained in solid-state diffusion processes at the surface of a silicon crystal lend themselves ideally to the fabrication of high-frequency transistors, operable at frequencies up to and above 1GHz.

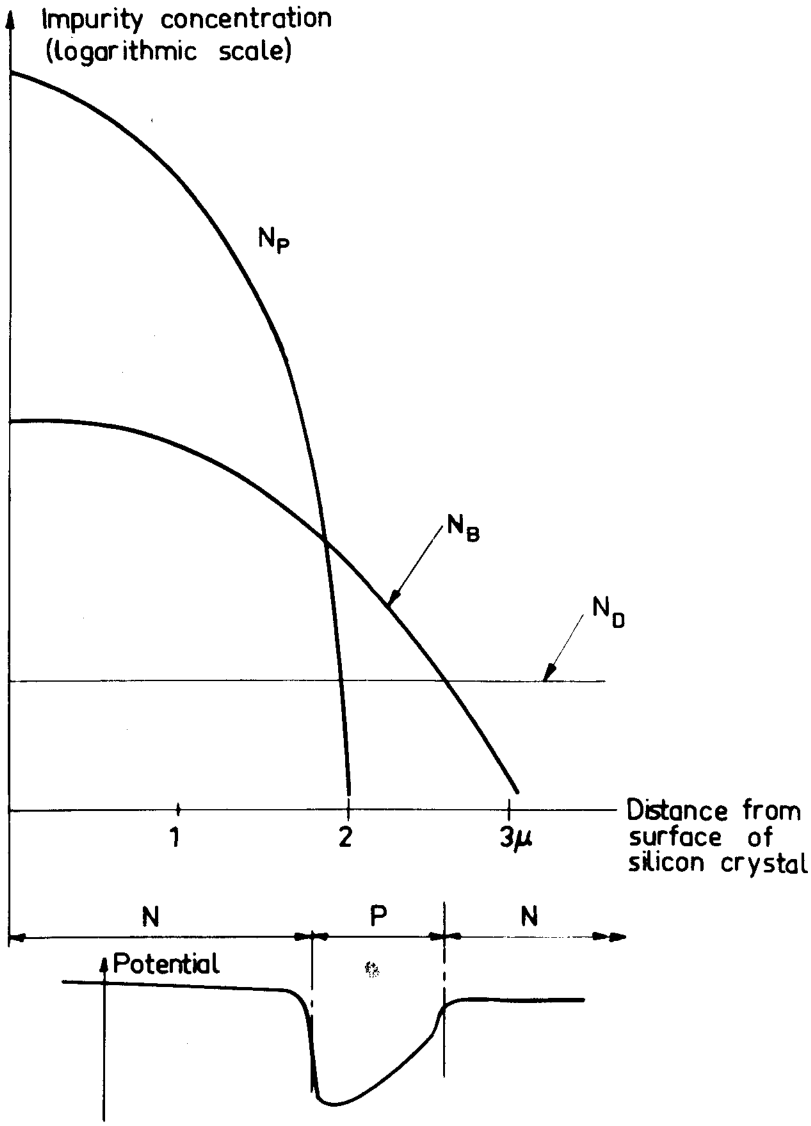


FIGURE 8.1 : CONCENTRATION OF BORON AND PHOSPHOROUS ATOMS AFTER TWO DIFFUSION PROCESSES AND DISTRIBUTION OF POTENTIAL.

8.3 **Oxide masking**

The discussion of solid-state diffusion in the previous section has shown that desirable distributions in depth of donor and acceptor impurities may be achieved in the surface layers of a silicon crystal for the purpose of fabricating a junction transistor. However there remain two problems: making electrical contacts to the three regions (particularly the p-type base region of Fig. 8.1), and limiting the active area of the device to a useful value.

An additional discovery in process technology enabled both these problems to be surmounted simultaneously. It has been shown that a thin layer of silicon dioxide (SiO_2), formed as a glassy layer in intimate contact with a silicon crystal, can effectively mask the underlying silicon against in-diffusion of some donor and acceptor impurities.

Accordingly in the fabrication of an n-p-n silicon transistor, following the steps outlined in Fig. 8.2, a layer of SiO_2 of thickness approximately $0.5\,\mu$ is first formed on the crystal by oxidation in water vapour at temperatures around $1150^\circ C$. Subsequently "windows" are opened in this layer by a selective etching process, and the diffusion of boron is carried out as outlined in Section 8.2. The p-type region is now seen to be confined to the lateral dimensions of the window, by the masking

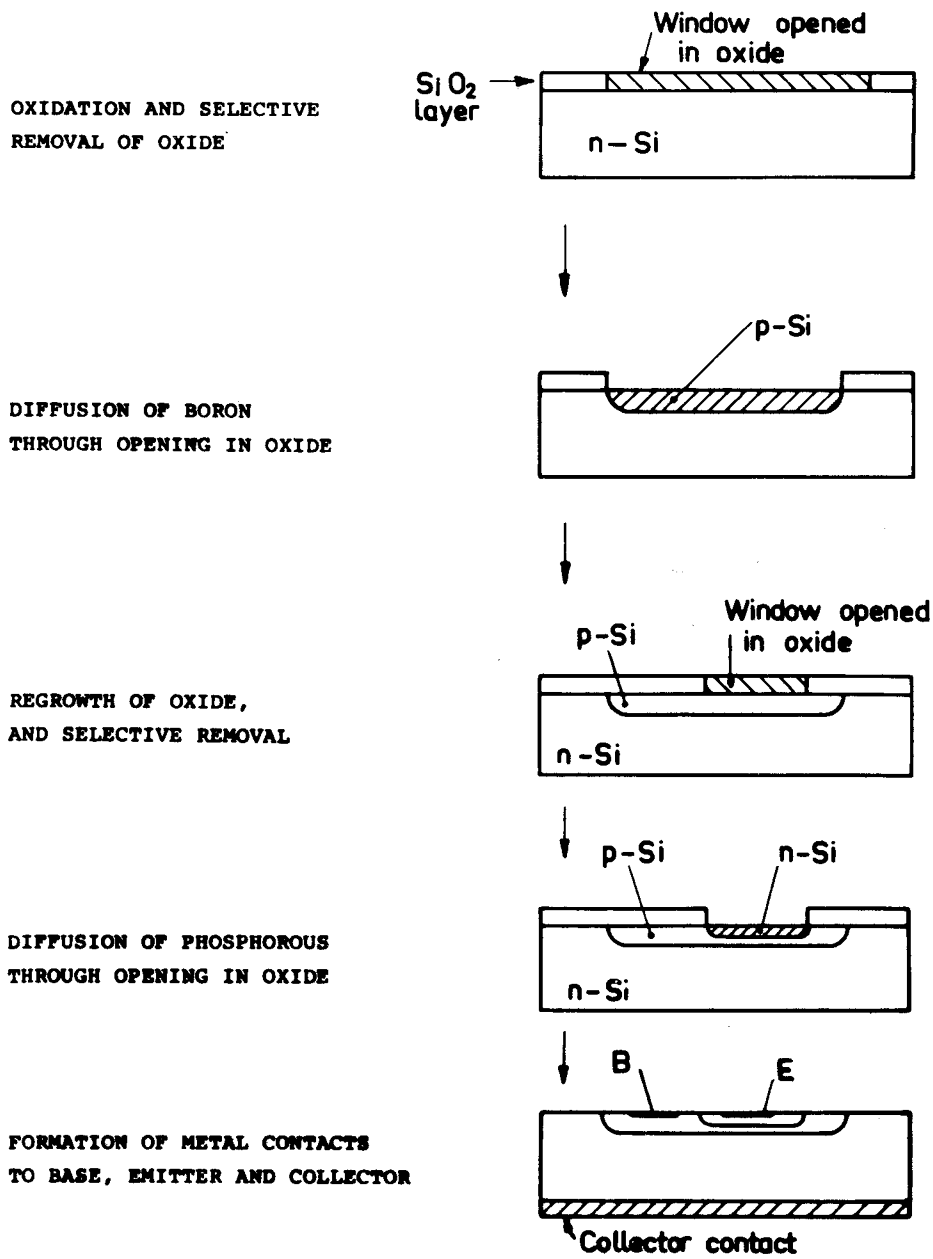


FIGURE 8.2 : STEPS IN THE FABRICATION OF A DIFFUSED SILICON NPN TRANSISTOR.

action of the SiO_2 layer. An oxide layer is then regrown; a second smaller window in this layer allows the diffusion of phosphorous to produce an n-type emitter region, again of limited lateral dimensions.

At this stage it is a relatively simple matter to make metallic contacts to the emitter and base regions at the top surface of the crystal; the collector contact can readily be made to the rear surface of the original n-type crystal. The crystal is broken into small segments, each containing one of the transistor structures of Fig. 8.2.

8.4 Planar technology

When a reverse bias voltage is applied to a silicon p-n junction the height of the potential barrier at the junction is increased, and with it the magnitude of the electric field in the depletion layer. As the reverse voltage is increased so also is the electric field. At field strengths around 100 kV cm^{-1} or above there is a dielectric breakdown in the junction: the reverse current increases markedly above the value I_S. (See Fig. 7.3).

Breakdown occurs because electrons and holes which traverse the depletion layer, contributing to I_S, are so strongly accelerated by the field that they become sufficiently energetic to break up some of the valence bonds. There is an avalanching increase of electron-hole pairs, and thus of the current carried by the increased number of charge-carriers. If the current is externally limited so that the junction temperature does not rise more than approximately $200^{\circ}C$, the breakdown may subsequently be extinguished without damage to the device.

At the surface which intersects a p-n junction there is always some electric field component in the space immediately adjacent to the silicon crystal. Charged ions originating from the ambient surroundings of the crystal may become absorbed on the crystal in the vicinity of the junction intersection, and in this situation will be influenced by the electric field associated with the junction. Their effect is to degrade the leakage current I_S, particularly in respect of noise content, and to lower the breakdown voltage below the value which would otherwise be achieved in the bulk of the crystal.

It is clearly desirable to protect the junction intersection from the ambient atmosphere, in order to avoid such degradation of the device properties, and a method of doing this is indicated in Fig. 8.3. Illustrated there is an enlarged view of a diffused p-type region, produced by diffusing boron through a window in an oxide layer on n-type silicon. It will be seen that there is a slight lateral diffusion of boron, as well as diffusion normal to the plane of the crystal surface, and that the intersection of the resulting p-n junction with the surface is actually protected from the ambient by the masking layer of oxide. To reinforce this protective action, it is customary for oxygen to be added to the surface atmosphere during diffusion so that a layer of borosilicate glass is formed on the silicon exposed in the window, as shown in Fig. 8.3. A contact window can subsequently be opened in this layer, well removed from the junction boundary, to obtain a p-n junction which exhibits a high degree of stability in its electrical characteristics with respect to variations in the ambient environment.

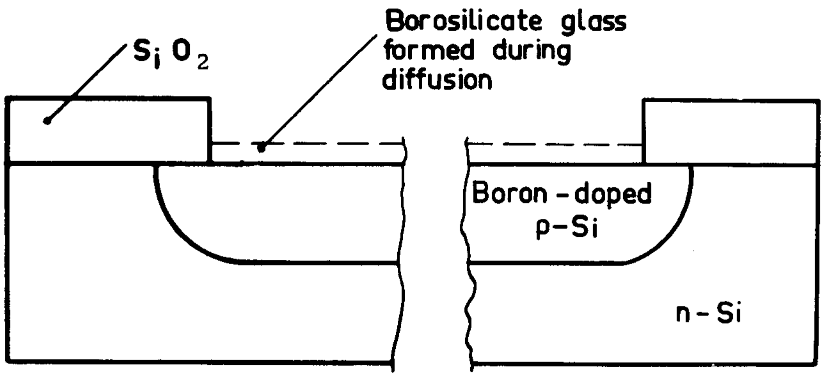


FIGURE 8.3 : PASSIVATION OF JUNCTION INTERSECTION WITH CRYSTAL SURFACE BY OXIDE LAYER.

8.5 Monolithic bipolar integrated circuits.

In order to open up windows in oxide layers for subsequent diffusion operations, extremely precise photolithographic techniques have been developed. The oxidized surface of a silicon crystal is coated with a photosensitive layer, which is polymerized and made resistant to acid when irradiated with ultra-violet light. When a photographic mask is placed in contact with this layer and exposed to UV radiation, regions of the mask which are blackened prevent the polymerization of the photosensitive layer immediately below them, and these non-polymerized regions can subsequently be removed with a solvent. A solution of hydrofluoric acid can then attack the oxide exposed through the windows in the photosensitive layer ("photoresist"), and produce windows of identical dimension in the oxide.

Optical techniques have been developed which enable windows of lateral dimensions as small as 2μ to be opened in SiO_2 layers, on a production basis. It is therefore possible to contemplate the manufacture of junction transistors whose dimensions are very much smaller indeed than the minimum size of silicon crystal chips which it is possible to handle mechanically (approximately 350μ square). Using one of the diffusion operations of the n-p-n transistor fabrication process outlined in the previous two Sections, it is also possible to produce diffused areas which can be used as resistors, as well as diffused junctions whose capacitance can serve as capacitors in a circuit, when reverse-biased. These possibilities lead naturally to the concept of a complete circuit diffused on and into the surface of an n-type silicon crystal. Required interconnections of components may be achieved by a pattern of metallization; a thin film of aluminium is evaporated continuously over the SiO_2 surface, and onto the silicon where contact windows have been opened, and is subsequently etched to the required pattern by a further photoresist stage.

Although the processes outlined above are complex, costs of integrated circuits produced in this way are kept low because of the large number (up to several thousand) that can be processed simultaneously on a single slice of silicon.

There is a number of design problems implicit in the technique outlined above for producing integrated circuits monolithically in a silicon crystal. Capacitances of p-n junctions, for example, are voltage-dependent, and there is a parasitic capacitance associated with diffused resistors. A further important problem, however, arises from the fact that all transistors have been diffused into an n-type silicon crystal, according to the procedure outlined above, and therefore have their collectors connected electrically within the crystal. This connection problem can be surmounted in the following way.

With the aid of a so-called epitaxial process it is possible to produce a monocrystalline layer of n-type silicon some 10μ thick on an underlying p-type substrate; the process is carried out at a temperature below the melting point of the p-type crystal, so that donor and acceptor impurities are kept distinct. By means of a preliminary boron diffusion through chosen areas of the epi-layer one is then able to produce "islands" of n-type material, as shown in Fig. 8.4, which are electrically isolated from each other by virtue of the low leakage current of reverse-biased silicon p-n junctions. Segments of an integrated circuit which it is desired to have non-connected electrically are then diffused in one or more of the islands produced in this way. We note that electrical connection to the collector of n-p-n transistors diffused in isolated regions must now be made at the surface of the crystal, as also shown in Fig. 8.4.

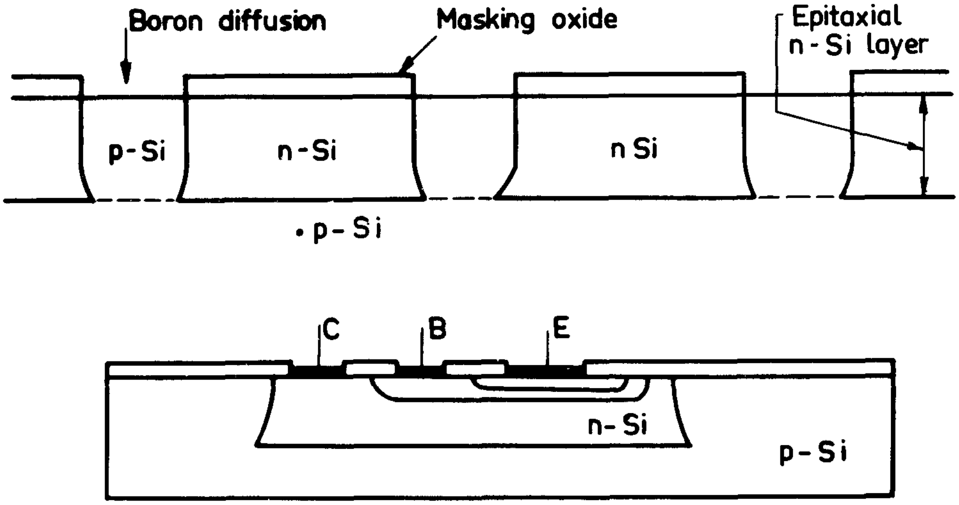


FIGURE 8.4 : ISOLATION, AND TRANSISTOR CONNECTIONS.

8.6 Conclusion

We have described in an elementary way the concept of a complete electronic circuit fabricated monolithically in a silicon crystal. A little thought will indicate that there are many problems, and many possibilities, associated with the design of such circuits which are not encountered in circuit design with more conventional discrete components.

The transistors discussed in relation to integrated circuit fabrication operate, as explained in Chapter 7, by the movement of carriers in excess of equilibrium from emitter to collector. The injected carriers – electrons in the base of an n-p-n transistor – are always accompanied by an approximately equal density of carriers of the opposite sign, in order to preserve space-charge neutrality, and for this reason such transistors are referred to as bipolar. Integrated circuits incorporating such transistors are also described as bipolar.

There is another class of transistor, a so-called field-effect transistor (FET), in which there are no injected carriers. Unfortunately there is no opportunity here to discuss the details of their operation, but it may be noted that significant advances in miniaturization may be achieved in the future with FETs, relative to bipolar integrated circuits.

8.7 Further reading

W.C. Hittinger and M. Sparks: "Microelectronics",
Scientific American 213, No. 5, p. 56, Nov. 1965.

The Western Electric Engineer 11, No. 4, Dec. 1967.

CHAPTER 9

BUILDING BLOCKS FOR SYSTEMS

"The world should know by this time that
one cannot reach Parnassus except by flying thither"

(from the notebook of G. M. Hopkins,
Oxford, April, 1864)

Chapter 9

BUILDING BLOCKS FOR SYSTEMS

9.1 Introduction:

In describing the nature of the building blocks from which descriptions of systems both large and small may be built up, it will be convenient to describe building blocks as if they were electrical or electromagnestic in nature. There is a fundamental reason for this choice, which is really a desire on our part to follow a pattern of development (i.e. of the ideas of system, system analysis and system design) which is similar to the pattern by which, historically, such ideas were formulated and put to use for mankind.

It is our intention to point out some of the mathematical descriptions which are used by the group of professional people nowadays known as "systems engineers". Originally developed mainly within the discipline of electrical engineering, these very general formulations are being found useful in many other fields of engineering and applied science including the life sciences. They are finding widespread application too in the social, political and military sciences. Already extensively used in administration and modern business management, they may also secure a place in the professional armoury of the educator and the humanist.

It is easy to see, therefore, that we must strive towards an extremely *generalized description* (i.e. a generally applicable mathematical model) for our building blocks in order that the *same sort* of description may be used profitably and easily for the extremely varied building blocks that we may encounter in all these different kinds of "systems". The way in which this can best be done is to seek out the simplest mathematical model capable of yielding the sort of information which we require.

This generalized but simple description was achieved in the electrical engineering field by three main steps

(i) a wise choice of variables,

(ii) by linearization of the mathematical models chosen to describe the building blocks (i.e. by discarding the complexities of non-linear behaviour),

(iii) by choosing to use what is known as the "black box" concept to describe the behaviour of a number of building blocks which have been connected together.

We will say more about each of these wise choices in what follows.

9.2 What Is a System?

In an earlier chapter, one of the suggested references is Machol "Systems Theory Handbook", in which is listed some of the attributes of a system. These are expressed in a way that is particular to that writer. If one looks at other authors one will see other expressions of what a system is and also of what a system is not. Probably not one of these catalogues, by itself, is a complete one. It is only by wide reading, by considering many different mathematical models and by working out a variety of examples that one acquires a firm notion of what is meant by the deceptively simple word; "system".

Quite often when one is reading or talking on this subject it is not even clearly stated whether the "system" under discussion is the actual physical collection of apparatus, devices, instruments, machinery and control knobs, levers etc., or whether the discussion refers to the mathematical model which has been built up in our minds and by using symbols written on paper in order to *represent* the actual physical system.

When one speaks of the behaviour of the system, one may mean either the actual behaviour of the system (i.e. speed, energy, position etc. at known times) or one may mean the algebraic or arithmetical behaviour (i.e. the manner of variation of the relevant variables concerned) of the mathematical model which has been conjured up or chosen to represent the actual system. Borrowing from the everyday jargon of computer engineers we might speak of the actual physical system as the "hardware" and the mathematical model as the "software".*

We will not attempt to give a canonical** definition of a system, but will proceed by way of examples and discussion to build up ideas as to what it is and what it is not.

* In the jargon of the computer world the word "software" is commonly used to mean a set of instructions for the computer i.e. a program or a compiler. Rather than being written as a set of symbols on a piece of paper, these might be "stored" as alpha-numeric symbols on punched cards, in magnetic-core memory stores, on magnetic tape or even in some of the newer computer-graphic systems in visual or photographic or holographic form. A mathematical model could be recorded in this way too.

** The word canonical occurs in mathematical terminology. Roughly speaking, it indicates the simplest possible form or description which is nevertheless complete or unambiguous.

9.3 Some Kinds of Systems

Let us list a few kinds of systems:–

Electrical systems, e.g. a large electric power supply system, a widespread telephone system, an electronic digital computer.

Mechanical systems, e.g. a steel rolling mill.

Chemical systems, e.g. an oil refinery.

Economic systems, e.g. government finance, industrial and commercial firms, personal monetary arrangements.

Biological systems, e.g. a mink farm, the tuna fishing grounds, an afforestation plantation.

Irrigation systems, e.g. the Aswan High Dam, the Mekong Delta scheme, the Murrumbidgee Irrigation Area.

Transportation systems, e.g. a state-wide railway system, the escalators in a large department store.

Biochemical systems e.g. the life support system in a space vehicle.

The list is a long way from being exhaustive; it should be easy for the reader to add other examples.

Not all of these systems may *best* be analysed in exactly the same way – however there are certain basic principles which may be applied in all cases to the models which we set up to describe their behaviour. Certainly, too, the methods developed in the last few decades for the analysis and design of electrical systems can be applied to many other kinds of dynamic systems.

The first common attribute of all the systems mentioned is that they are dynamic. By the word *dynamic,* is meant a system whose *state* is *varying with time.*

The state of a system may be described by one or more variables:– for example, in the telephone system we might be interested in the *number of conversations* being carried on *at a given moment* and to describe this we need to symbolize the number of conversations as a variable function of time; thus we might choose the symbol n_1 (t) to represent this quantity at each specified instant of time t. However we might also be interested in the number of *trunk line* channels *occupied at a given moment* and we would then need another symbol to represent this quantity, say n_2 (t). Again for example, in the life support system of a space capsule we would certainly be interested in the pressure, temperature, percentages of oxygen, carbon dioxide and other compounds and we could write each of these variables as a function of time in order that we might study the dynamics of the system i.e. the way in which these variables change as time progresses.

We have taken our first step towards a mathematical model by recognizing that we are concerned with system dynamics and that it is thus desirable to describe the state of the system by variables which are functions of time. Let us note at this stage that some of these variables may be independent variables and others, dependent variables.

9.4 Conversation Laws or Principles

Common to the problem of describing *all* sorts of systems is the notion of choosing a *suitable conservation law* or *principle* as a basis for our mathematical model.

Examples of these are:–

In a *chemical* system – conservation of *mass*

In an *electrical* system – conservation of *charge*

In a *combined* system (e.g. electro-mechanical and thermal)
– conservation of *energy*

In a *closed economic* system – conservation of *money*

We must realize that any one of these conservation laws will *hold only in* a suitably *closed* and defined *system.* Whenever we have to extend beyond our previously closed system, we may have to modify the conservation laws.

For example, in an atomic explosion the two everyday laws of conservation of energy and conservation of mass are both violated – some mass disappears and a large amount of energy appears. In this case the modification is to take account of $E = mc^2$ and to say that (Mass + Energy) is conserved.

In addition we must also know all imports and exports to and from our closed system and include these in the *balance* or *conservation equation.* Let us commence with a homely example.

We all know the phrase to *balance one's budget.* Most of us have been faced with this problem in the past and no doubt most of us will continue to be faced with the same problem for the whole of our lives. In essence the law of *conservation of money** may be expressed by the equation.

$$\text{Income} - \text{Expenditure} + \text{Stored Money} = 0$$

When expenditure exceeds the sum of income and stored money we have to go *outside* our closed and *borrow* (assuming we do not wish to steal) from friends, family or banker. This will mean that we need an extra term in the equation i.e. Borrowings.

Unfortunately the act of borrowing (unlike stealing or begging alms) usually means that we must include the further term *Repayments* and very often too, *Interest.* The equation of conservation of money now looks like

$$\begin{aligned}&\text{Income} + \text{Borrowings}\\ &\quad - \text{Expenditure} - \text{Repayments} - \text{Interest}\\ &\quad\quad + \text{Stored Money} = 0\end{aligned}$$

It is this equation (possibly with extra terms to allow for gifts, acts of god such as wrecks at sea or destruction by fire etc.) that accountants have used to set up daily weekly, monthly or annual accounting systems.

It should be observed that all these terms would normally be functions of time.

It should also be noted that there could be some difference of opinion about the correct way to specify what is meant by stored money – if we are considering a time interval such as a week or a month we should really write down the *change in stored money* over this interval of time. The next example but one may make this clearer. Also to be noticed is the adoption of a negative sign for this term and the reader should make sure he understands why this is so written.

Another way of expressing a conservation law is to say (as in the conservation of mass in ordinary chemical reactions)

$$(\text{Total mass involved}) = \text{Constant}$$

There are two ways in which the last equation may be rewritten to get a zero on the R.H.S.**

(1) $(\text{Total mass at start}) - (\text{Total mass at finish}) = 0$

OR, (2) $\frac{d}{dt}(\text{Total mass involved}) = 0$

and both these forms are useful.

Form (1) holds for a finite time interval

Form (2) is an instantaneous balance.

Other laws may be written down as examples e.g. conservation of energy.

$$\begin{pmatrix}\text{Total Energy}\\ \text{Input}\end{pmatrix} - \begin{pmatrix}\text{Total Energy}\\ \text{Output}\end{pmatrix} - \begin{pmatrix}\text{Increase in}\\ \text{Stored}\\ \text{Energy}\end{pmatrix} = 0$$

This equation would refer to a closed system with inputs, outputs and capability of storage. It refers also to a defined time interval.

It is well known that $\frac{d}{dt}(\text{Energy}) = \text{Power}.$

So by differentiating the above conservation of energy equation we would get

$$\begin{pmatrix}\text{Input}\\ \text{Power}\end{pmatrix} - \begin{pmatrix}\text{Output}\\ \text{Power}\end{pmatrix} - \begin{pmatrix}\text{Time Rate of}\\ \text{increase of}\\ \text{stored energy}\end{pmatrix} = 0$$

* This is not a generally recognized name for this relationship. It is used here to draw the analogy with other systems.
** This is a convenient but not essential way in which to write down some equations.

This is also a conservation equation but it refers to *instantaneous conditions* in the closed system with specified input and output arrangements. It is therefore a more convenient formulation when we wish to express the three quantities

Input, Output and Storage

as functions of the variable, time.

In a biological system (it may be more correct to call this example ecological) we might have for the total population of a closed system (such as e.g. the human population of Australia, or the tuna population of certain fishing grounds off the coast) a conservation equation such as

$$\begin{array}{lcl} \text{(Immigrants)} & - & \text{(Emigrants)} \\ + \text{(Births)} & - & \text{(Deaths)} \\ \quad - \text{(Increase in population)} & = & 0 \end{array}$$

(written for a specified time period)

Again by differentiating with respect to time would arise a similar equation = 0 but with each term representing a *rate* (hence the terms birth rate, death rate etc.)

From these seemingly obvious statements may be developed the *dynamics* of *systems.*

9.5 Input and Output

Let us now turn to a new use of the words *Input* and *Output.*

Suppose we have a dynamic system G

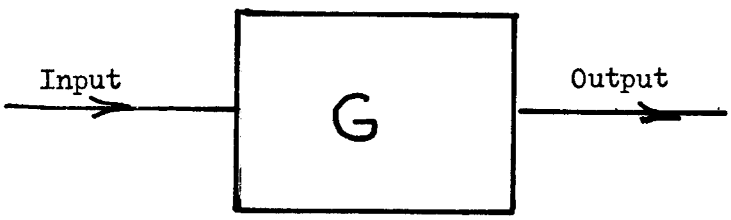


Fig. 9-1

We can consider the *system* receiving from the outside world a *certain input* which is a *function* of *time.* (e.g. consider the system comprising a motor car plus driver on a straight open road, whose *input* is the position of the accelerator pedal as a function of time).

We can also consider the system as giving to the outside world a *certain output* which is also a *function of time* (e.g. the *distance* of the *car* from some specified starting point).

Note 1: we could have specified *other* inputs and outputs. This is quite arbitrary and our choice simply indicates what we are interested in.

Note 2: we could have specified two inputs (e.g. *accelerator* pedal position and *gear* change)

Note 3: we could have specified two or even three outputs in which we are interested (e.g. *distance* covered, *speed, acceleration).*

In the latter case we need additional equations or we can if we wish use arrays of numbers and the rules of matrix algebra in order to calculate all the outputs *simultaneously.* For those readers who have not yet studied matrices it may be sufficient to say that matrices look like determinants, and are treated in more advanced texts. An array of numbers (or symbols) is used to designate the inputs or outputs. These are single column or single row matrices and are often termed "vectors". They

should not be confused with directional vectors in real space. They are however vectors in the "space" formed by the two (or more) variables we are dealing with. Most of the readers of this book will have already encountered determinant arrays as a means of solving simultaneous equations. In this case most of the numbers in the array represent coefficients of terms in the equations. The solution of simultaneous equations is possible because there are certain formal rules for manipulating determinants (and matrices).

Now let us return to single input and single output systems.

9.6 Signals

It is part of the terminology of systems theory that we speak of input and output signals to a system. Basically these are simply variable quantities which are either known or else have to be calculated, as functions of time.

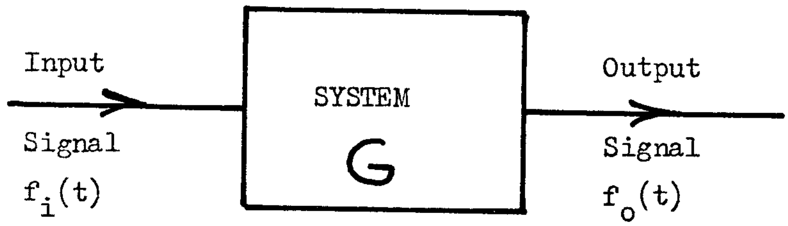


Fig. 9-2

Electrical engineering circuit theory and systems theory as we know it today is based on the *response* (i.e. output) of a *linear system* to a specified *excitation* (i.e. input).

Both response and excitation are *functions of time.*

The theory may be extended also to *non-linear* systems* but these are considerably *more complicated* and also very much less well understood.

The *linear* theory is based on the mathematical analysis of *linear ordinary differential equations.* You have all dealt with some examples of ordinary differentials equations in your Maths courses at school. Some of you may also have realized that differential equations in general are divided into two classes

(i) ordinary differential equations

(ii) partial differential equations

In each of these classes we may have either linear or non-linear cases. We will say a little more about partial differential equations when we come to distributed systems. For the moment let us consider only the systems describable by ordinary differential equations. These may be recognized by the fact that the differential operators are written in the familiar form $\frac{d}{dt}, \frac{d}{dx}, \frac{d^2}{dt^2}$ etc. while partial differential operators are written as $\frac{\delta}{\delta t}, \frac{\delta}{\delta x}, \frac{\delta^2}{\delta t^2}$ etc. or by making use of the symbol ∇ ' (pronounced del or occasionally nabla) of the branch of algebra known as vector analysis.

A complicated system may be represented by a set of several simultaneous differential equations. We will not use examples of this complexity – the necessary concepts may be conveyed more easily by considering examples where the system is representable by a single differential equation. The solution of such a single equation is usually found in two parts. To be specific , let y be an unknown variable and let x (or t) be a known variable.

* See e.g. W.J. Cunningham "Non-linear Analysis" (McGraw-Hill 1958).

Let us write the differential equation in the following standard format

$$\underset{\text{L.H.S.}}{\begin{pmatrix} \text{Collect terms in y and its derivatives} \\ y,\ \frac{dy}{dx},\ \frac{d^2y}{dx^2}\ (\text{or } \frac{dy}{dt})\ \text{etc.} \end{pmatrix}} = \underset{\text{R.H.S.}}{\begin{pmatrix} \text{Collect terms} \\ \text{in x} \end{pmatrix}}$$

In other presentations the variable x may be replaced by time. The two parts of the solution are

(1) The *complementary function*, obtained by putting zero in place of the R.H.S.

This part is often termed the *TRANSIENT solution* or *free response.*

(2) The *particular integral*, when the R.H.S. is the (particular) known function (often called the forcing function or driving function). This part is often termed the *STEADY STATE solution* or *forced response.*

Notice that the forcing function is the *input signal* of Fig. 9-2, while the complete solution (C.F. plus P.I.) is the *output signal* of Fig. 9-2.

A typical response to a sinusoidal input signal switched on at time t = 0 is shown below

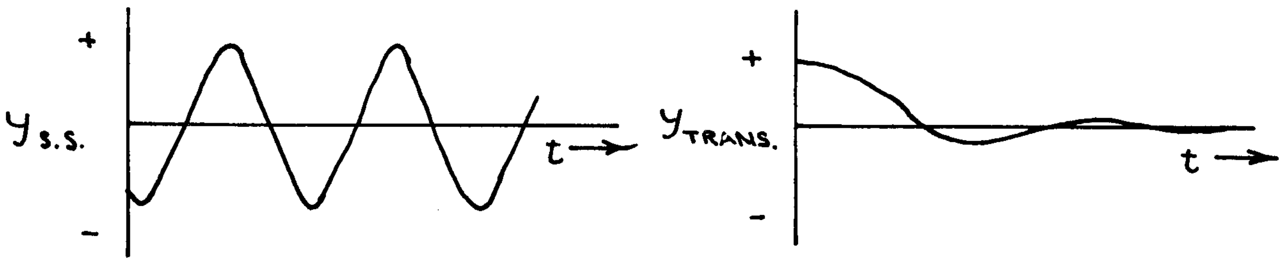


Fig. 9-3

The complete solution is $y = y_{TRANS} + y_{SS}$
and each of these three quantities is a function of time.

In a well behaved (i.e. stable) system the *transient dies out* while the *steady state* response keeps on as long as the input signal is present.

9.7 Methods Available To Calculate The Response

We will return in the next chapter to look a little more closely at the methods available to calculate the response of a system to a specified input. Other texts used in later years of undergraduate courses study each of these methods in detail.

Readers should learn more of the advantages and disadvantages of each scheme (and the variations of each scheme) by undertaking a good deal of problem solving. Some parts of this task can be quite tedious, but it is worthwhile to reflect that it is not much use just knowing how to catch fish – if life is to be sustained we must also practice the occupation of fishing and, what is more, we must be able successfully to catch actual fish, not just snags or previously discarded rubbish.

To conclude this chapter let us say something about a number of actual input/output devices. That is, let us look briefly at some hardware.

9.8 Amplifiers, Machines and Transducers

It would have been possible to add a great many more nouns to those chosen for the above subheading. Notice that each of the three nouns describes an *input/output system.* The reader should try to make up his own list of nouns each of which describes a particular kind of input/output system. Even if we confine the list to manmade hardware systems by excluding

biological, social, sexual, political, managerial, intellectual, emotional, educational etc. etc. systems, we will still quickly compile a very long list.

It is quite natural to ask – how is it that we can describe these many different things by the one word "system". Is it just an omnibus word, a dust bin of sorts? Or can we find attributes which are common to all "systems"? The answer to the last question will turn out to be "Yes", but first of all let us look at a few particular devices, remember that while each one is a small system just by itself, and therefore worthy of attention, nevertheless that our real aim is to gain understanding of how to connect together a number of devices in order to form larger systems.

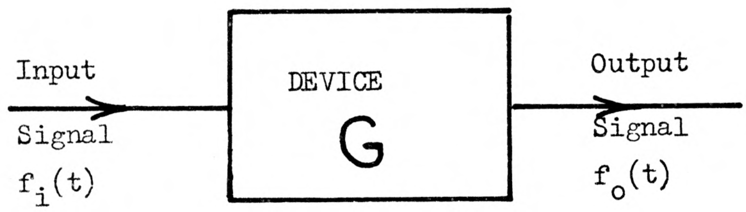


Fig. 9-4

The sketch in Fig. 9-4 (which is almost but not quite a repeat of Fig. 9-2) can represent an amplifier, by which we mean (usually) a device which when provided with an input signal will produce a similar but larger (i.e. amplified) output signal. In an ideal amplifier the output will be a replica of the input, produced instantaneously and with increased amplitude. We may write for the response of such an ideal amplifier, using subscripts "o" for output, "i" for input

$$f_o(t) \quad = \quad G\ f_i(t)$$

where G is a constant real number greater than unity. In practical amplifiers there will be some difference in the shape of $f_o(t)$ and $f_i(t)$ in addition to the difference in amplitude described by a constant real G.

This difference represents distortion and is described in a variety of ways. Which way we choose to describe distortion depends on the manner in which the input signal itself is described. Thus G may be complex, rather than purely real; G could be a function of the input signal; or we could allow additional classes or modes of signal to be present in $f_o(t)$ over and above those which are present in the input $f_i(t)$*.

Amplifiers in which the input and output are electrical signals are usually made up with electronic *active devices* (i.e. controllable amplifying devicesas distinct from the *passive* electrical components such as resistance, inductance and capacitance). However in addition to the various kinds of vacuum tube and transistor, active devices may be made up by utilizing the non-linear properties of ferro-magnetic cores and of some dielectric materials. In addition it is possible to utilize the non-linear properties of electric conduction through ionized gases. A device which behaves very much like a transistor (but is much slower) can be made up using ionic conduction in liquids and semi-permeable membranes. No technological use has been found for the latter device but a study of its behaviour could serve to give extra insight into the mechanisms of the periphereal nervous system in animals and humans.

Other forms of amplifier are the controlled switch either in the form of the electro-mechanical relay, or the controlled gas-discharge tubes known as thyratrons and ignitrons, or in more modern guise the silicon-controlled-rectifier (SCR) or thyristor. The relay contacts are *controlled* (i.e. opened or closed) by action of current in an electromagnet. Conduction in the gas-discharge tubes is initiated by an electrical impulse applied to a grid or igniter electrode. In the S.C.R. conduction is initiated by an impulse applied to a small gate electrode (actually a subsidiary p-n junction) built into or onto the main p-n junction. In the latter two cases conduction is quenched (or commutated, if it is a repetitive on-off operation) by allowing the current to fall to zero for a sufficient time or by superimposing a larger current pulse in the reverse direction.

* A simple example of this occurs when a sinusoid input generates harmonic frequencies which occur in the output i.e. signals at twice, three, four etc. times the frequency of the original input or so-called fundamental frequency. A slightly more complicated case is when the input is the sum of two (or more) sinusoids of differing frequency. Commonly occurring distortion modes occur at the sum or difference frequencies of the various fundamental and harmonic frequencies. For a square step wave input the distortion is usually described by referring to the finite rise-time of the output and its degree of "overshoot" before settling down to its new steady value. For certain pulsed waveforms as may occur in a television signal it is convenient to describe the distortion in terms of paired echoes (Reference Goldman S. Signals Modulation and Noise (McGraw Hill, 1948)) or in modern television practice in terms dictated by convenient and accurate measurement techniques.

Such *switched-mode amplifiers* are capable of extremely high power gains. The power gain may be defined as the ratio of "power controlled" to "power needed to operate the switch". Power gains between 10^2 and 10^5 are readily achieved.

Switched-mode amplifiers may also be built up using transistors or vacuum tubes. For example the so-called Class "C" amplifier used as the final or power stage of a broadcast or T.V. transmitter operates in a switched mode and may have a power gain approaching 10^2. As another example the so-called flip-flop circuit using transistors is really a two way switch which changes state in response to an impulse input. This device finds very wide application in digital logic circuits (see Chapter 13)..

In switched-mode devices we usually exploit the high energy efficiency made possible by having the device either fully ON or completely OFF.

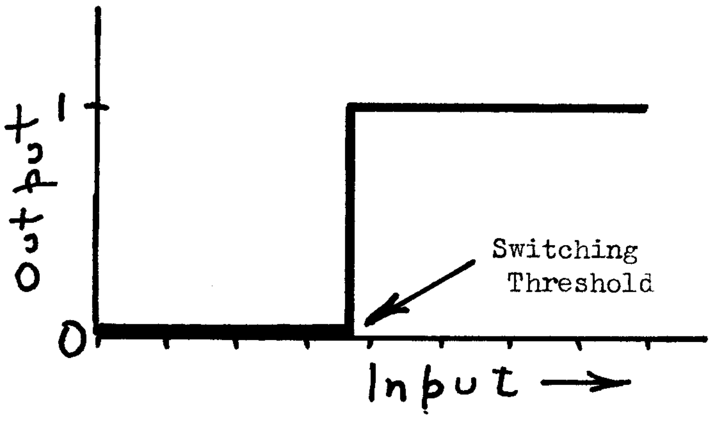


Fig. 9-5

Fig. 9-5 illustrates the "characteristic" or input-output behaviour of a simple switched device. It is easy to see that G in the expression

$$f_o(t) \quad = \quad G.\, f_i(t)$$

is a non-linear function which is zero below the threshold and such that the product $G.f_i(t)$ is unity above the threshold.

In distinction to the switched-mode amplifiers is the class of *linear amplifiers*. In these devices we endeavour to make the change in output proportional to the input. A characteristic for this class of amplifier is shown in Fig. 9-6. The relation between output and input is a linear function.

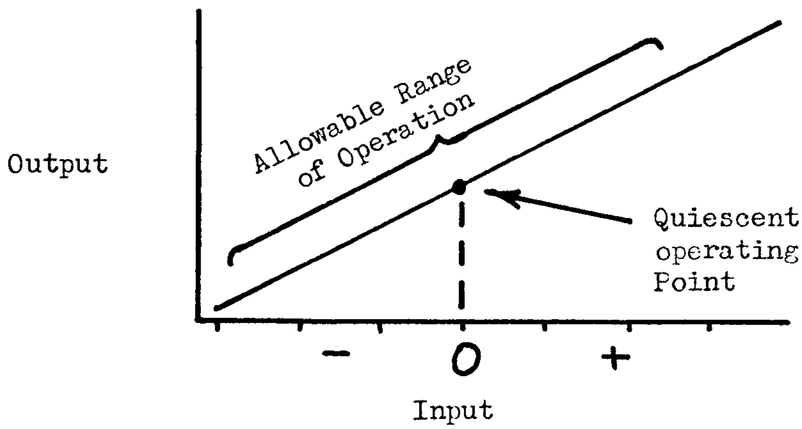


Fig. 9-6

By connecting two amplifying devices in a so-called "push-pull" connection the operating point may be established at the origin, as is suggested in Fig. 9-7. This artifice reduces distortion and increases the allowable operating range.

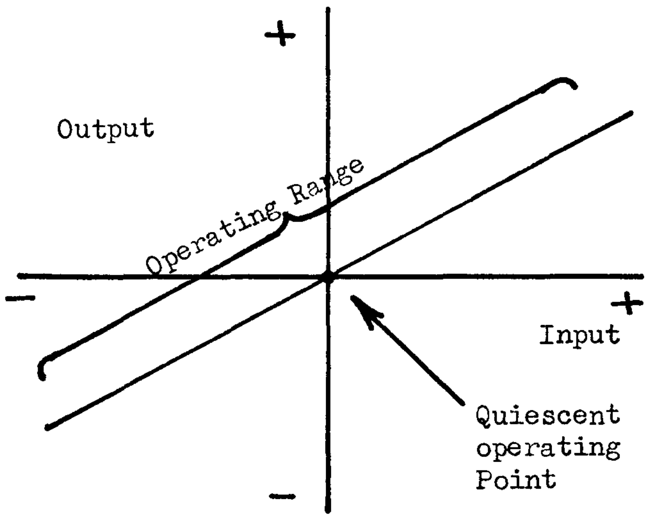


Fig. 9-7

The word quiescent is used to indicate the condition of the amplifier at zero-input.

Clearly switched-mode and linear-mode amplifying are two extremes and in that sense are ideal. Practical amplifiers will have characteristics which depart to some extent from those shown.

In addition to electronic or electrical amplifiers, other forms of switched-mode and linear amplifier are possible. Among these are fluidic devices in which both the stream-line and turbulent flow properties of a fluid are exploited – by the expenditure of a small amount of input flow or pressure energy a much larger amount of output flow or pressure may be controlled. The devices commercially available are particularly suitable for switching-mode operation and fluidic logic systems are increasingly coming into use. Air if used as the working fluid has the advantage of much lower inertia than a liquid with consequent reasonably rapid response even from a multi-device system.

Hydraulic amplifiers have been used since the last century. The mechanical forces developed by quite small devices may be made very large and by careful design of the control valves to minimize the controlling force needed (e.g. by a balanced arrangement) large ratios of power gain may be achieved. Typical applications include the raising and lowering of aircraft landing wheels, the controlled movement of large masses such as guns or gun turrets, the operation of certain powered auxiliaries in motor vehicles etc.

Without having specifically defined what we mean by the terms, we have already cited some examples of machines and transducers. It would not be easy to give a satisfactory definition of what we mean by a machine. Of the various likely attributes of a machine, which ones should we emphasize? Some of these likely attributes include – it is man-made, it involves mechanical forces and movements, it consumes or delivers or stores mechanical energy.

A transducer is more readily defined – its meaning has not become diffuse since it has not yet passed into everyday non-technical use nor has it (yet) come into the hands of the advertising copy-writers. A transducer is simply a device for the conversion of energy from one form into another. The input and output can both be expressed in energy terms but different kinds or forms of energy. Examples are given in the table 9-1.

Table 9-1

Transducer	Input Energy	Output Energy
Loudspeaker	Electrical	Acoustic
Microphone	Acoustic	Electrical
Electric lamp	Electrical	Luminous and infra red
Photo-cell	Light	Electrical
Air compressor	Mechanical	Compressed gas
Electric generator	Mechanical	Electrical
Steam engine	Chemical	Mechanical
Petrol engine	Chemical	Mechanical

It will be noticed that in every case the transducer may be represented by the block diagram of Fig. 9-4 or 9-2.

Individual devices may be arranged so that an output of one serves as an input to another device. We will learn more about the ways in which this may be done in the next chapter.

9.9 Summary of Highlights in Chapter 9

(i) A system possesses inputs and outputs, but is otherwise isolated.

(ii) The outputs depend on the inputs and the initial state of the system.

(iii) The behaviour of a system may be described by a set of equations. These are known as the mathematical model of the system.

(iv) The choice of variables is important and will decide whether the mathematical model is a useful one.

(v) Inputs and outputs vary with time (i.e. are functions of time) and are often termed signals.

(vi) The set of equations is usually derived by applying suitable conservation laws to the system.

(vii) The closed system is represented conventionally by a closed box with one or more inputs and one or more outputs e.g. Fig. 9-2.

(viii) Amplifiers, machines and transducers are important devices because they serve as building blocks for systems.

(ix) Amplifiers, machines and transducers may operate either in a switched-mode or in a linear mode.

(x) In the representation or model of a system, each device may be represented as a block.

(xi) Individual blocks may be connected together to form a larger system, which could in turn be represented by a single block if we so wished.

(xii) We will learn more in Chapter 10 about the ways in which devices (or the blocks representing them) may be connected together to form larger systems. We will also learn something of the way in which a mathematical model for a system may be formed provided we know the properties of and the connections between the component blocks.

9.10 Exercises

1. Make a list of further examples to add to Table 9-1.

2. Draw a block diagram for a small record-player. Which blocks represent transducers?

3. Draw a block diagram for a thermostat controlling the temperature of an electric oven. (The thermostat is a switch). Draw a block diagram for a thermostat controlling the temperature of a gas oven. (The thermostat regulates the flow of gas, but it is not a "switch").

9.11 Suggested Reading Material

M.A. Lynch and J.G. Truxal, "Introductory System Analysis" (McGraw-Hill 1961) – Preface, Chap. 1 and Chap. 3 up to and including Section 3.7

Chapter 10

NETWORKS OF BUILDING BLOCKS: SYSTEMS

"The Hunter with his dogs pursues his circuit.
. . . . . The endless cycle of idea and action,
Endless invention, endless experiment
Brings knowledge of motion, ... "

(First chorus from "The Rock", T.S. Eliot)

Chapter 10

NETWORKS OF BUILDING BLOCKS: SYSTEMS

10.1 Introduction

It is sometimes quite hard to tell, when using the word system, whether we are referring to the actual system itself or whether we are referring to the model of the system, which has been built up in our minds. In this chapter (and subsequent ones) the reader should try consciously to ask and to answer this question – are we referring to the actual system or to its model? Sometimes the distinction may be important, sometimes it may not matter.

This chapter and the one following deal in simple terms with some of the mathematical models which have been found useful for quantitative prediction of the behaviour of systems. The same mathematical models have been found useful for the other task of designing systems which will exhibit a desired behaviour.

10.2 Methods Available To Calculate The Response Of A System

The choice of the most suitable method of calculation will depend on two things

(1) the kind of signal – the word "mode" is sometimes associated with this

and (2) the way in which a mathematical model for the system has been built up.

We may say that the three things – signal, system and computational method must be compatible i.e. they must all three fit together reasonably neatly.

Let us consider two examples.

10.2.1 FIRST EXAMPLE – which is concerned only with steady state response

First, when the input signal is a sinusoid* $E_i \cos \omega t$ and we may use the electrical engineer's idea of *impedances* or phasor algebra in order to obtain the response of the system. This is described and taught as A.C.** circuit theory, and may be presented in undergraduate courses in an early Physics course and carried on in Electrical Engineering middle year courses. The basic signal of phasor algebra is the sinusoid of pre-specified frequency. It is an important point of the theory that the *steady state response* of a *linear system* to a sinusoidal input is another sinusoid of *identical frequency*.

Let us examine a block diagram which suggests the mathematical manipulation needed to obtain the response of a linear system to a sinusoidal input.

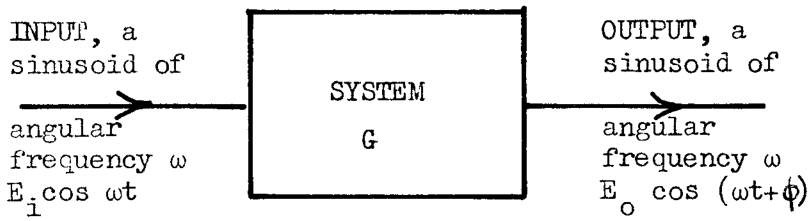


Fig. 10-1

* Both $E \cos \omega t$ (where E and ω are constants) are described as sinusoids in the variable t, or sinusoidal in time. So also is the more general expression $E \cos (\omega t+\phi)$ called a sinusoid where ϕ is another constant known as the (initial i.e. t = o) phase angle. E is amplitude while ω is angular frequency in radians per second if t is measured in seconds. Also $\omega = 2\pi f$ where f is frequency in hertz or cycles per second.

** Alternating current, by which is implied the particular kind of alternation described by a sinusoid of *constant* amplitude, frequency and phase angle.

This is the same as the block diagram Fig. 9-2 or Fig. 9-3, *except* that we have *specified* the exact form or mode of the input and output signals. In this particular case the specification we have adopted for the signals has implied that we are concerned with (and *only* with) the *steady-state* portion of the response. By this limitation we have also limited the way in which the system function G can (and indeed must) be written down.

Consider that we wish* to be able to use a relation

$$\text{(Output)} = \text{(System Function)}\,\text{(Input)} \qquad \ldots.(10\text{-}1)$$

in order to calculate the response to a specified input. In operational terms this means

$$\text{(Output)} = \begin{array}{l}\text{result of operating on the (Input)}\\ \text{with the operator (System Function)}\end{array} \qquad \ldots.(10\text{-}2)$$

and we also wish to be able to carry out the *operation* using the ordinary rules for algebraic *multiplication.* It will be shown that this is possible, *provided* that each of the three quantities (Output), (Response) and (Input) are written as complex numbers of a kind known as phasors**

Multiplication of these phasors is expressed in a phasor equation i.e.

$$\begin{pmatrix}\text{Output}\\ \text{Phasor}\end{pmatrix} = \begin{pmatrix}\text{System}\\ \text{Function}\\ \text{Phasor}\end{pmatrix}\begin{pmatrix}\text{Input}\\ \text{Phasor}\end{pmatrix} \qquad \ldots.(10\text{-}3)$$

The two operations on the right hand side which are needed to obtain the answer for the left hand side are simply

(i) Multiply the magnitudes

and (ii) Add the phase angles with due regard to sign.

Having established the idea of using equation (10-3) in that form (i.e. for *analysis,* where output is the unknown quantity) we can turn it around to use for two other problems –

$$\begin{pmatrix}\text{System}\\ \text{Function}\\ \text{Phasor}\end{pmatrix} = \begin{pmatrix}\text{Output}\\ \text{Phasor}\end{pmatrix} \div \begin{pmatrix}\text{Input}\\ \text{Phasor}\end{pmatrix} \qquad \ldots.(10\text{-}4)$$

$$\begin{pmatrix}\text{Input}\\ \text{Phasor}\end{pmatrix} = \begin{pmatrix}\text{Output}\\ \text{Phasor}\end{pmatrix} \div \begin{pmatrix}\text{System}\\ \text{Function}\\ \text{Phasor}\end{pmatrix} \qquad \ldots.(10\text{-}5)$$

In each case we have written the unknown quantity on the left hand side. In equations (10-4) and (10-5) these calculations are both possible and easy because simple rules † can be made for phasor division.

The three equations concerned may be tabulated as shown in Table 10-1.

The identification problem usually refers to the process where output and input are measured on an actual system and the calculation is undertaken to *identify* the properties or coefficients of the system function. The synthesis problem is usually stated somewhat differently; given a specified input and a (specified) desired output, and *also* given a possible manner of connection of the components in the system, then let us calculate the magnitudes of the components involved. The synthesis

* The reason for this will be discussed a little later in this chapter.

** A phasor is a complex number usually expressed in polar form where the magnitude represents amplitude (base-line to peak) of a sinusoid and the angle represents the phase angle of the sinusoid. In the case of the phasor system function G of Fig. 10-1 the magnitude represents the ratio $\frac{E_o}{E_i}$, while the angle represents the shift in phase angle from (Input) to (Output). In Fig. 10-1 this shift in phase angle is ($+\phi$) radians. The output is obtained from the input by use of the two rules

Magnitude E_o = (Magnitude of G) times (Magnitude of E_i)

Output Phase Angle = Input Phase Angle + System Function Phase Angle.

† To divide one phasor by another we divide the magnitudes and subtract the phase angles.

problem is a *design* problem and in general does not have a unique answer since there may be many different ways of connecting up components which could achieve the desired result for a given system function. For one manner of connection (i.e. for a given circuit configuration) there is usually only one set of values for the components*, if we have specified the problem correctly.

A *reminder* may be useful before going on to the next section:–

Remember all our discussion in this chapter has referred and will refer to *linear systems.*

Table 10-1

Problem usually named as	Known	Unknown	Equation
Analysis	Input and System Function	**Output**	(10-3)
Identification Synthesis	Input and Output	System Function	(10-4)
Recovery OR Inverse Analysis	Output and System Function	Input	(10-5)

SECOND EXAMPLE – **where we consider both transient and steady state responses**

Both the steady state and transient responses may be obtained by the use of the Laplace Transform method. This scheme of calculation is more general than that of phasor algebra. It is based on the theory of functions of a complex variable which is well treated in the more advanced texts used for second and third years of engineering courses.

Let us draw a block diagram to represent what goes on in the Laplace Transform method for calculating system responses.

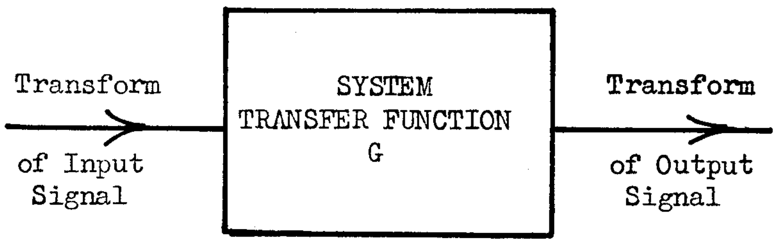


Fig. 10-2

* Provided we exclude negative values for the coefficients and provided that the configuration does not result in indeterminate results in the arithmetic. The latter can occur for example in certain bridge or tee configurations.

The similarity of this with Figs. 10-1, 9-2 and 9-3 should be noted. The meaning of Fig. 10-2 may be expressed by saying

$$\begin{pmatrix}\text{Transform}\\ \text{of Output}\end{pmatrix} = \begin{pmatrix}\text{System}\\ \text{Transfer}\\ \text{Function}\end{pmatrix}\begin{pmatrix}\text{Transform}\\ \text{of Input}\end{pmatrix} \qquad \text{....(10-6)}$$

or at somewhat greater length

$$\begin{pmatrix}\text{Output}\\ \text{Transform}\end{pmatrix} = \text{result of operating on the} \begin{pmatrix}\text{Input}\\ \text{Transform}\end{pmatrix}$$

$$\text{with the operator} \begin{pmatrix}\text{System}\\ \text{Transfer}\\ \text{Function}\end{pmatrix} \qquad \text{....(10-7)}$$

The basic signal or mode of time-variation which is associated with the Laplace transformation, is a *damped sinusoid**, or if we wish to build up more complicated cases a *summation* of *different damped sinusoids.*

Let us now try to gain some idea of the reason why the Laplace Transform method is so useful for system calculations. Consider the following analogy and work out also for yourself the way in which this analogy could be applied to illustrate the behaviour of phasor algebra as a simple operational means of solving (i.e. computing) what would otherwise be a more difficult mathematical problem.

The analogy goes as below:

Addition (or subtraction) is much *easier than multiplication* (or division).

Suppose we wish to multiply p by q. Let us transform each quantity by taking logs. The problem is now to calculate

log p + log q in order to get log pq

The operation or process of multiplication has been transformed into the simpler operation or process of addition.

The problem is completed by an inverse transformation being applied to log pq (i.e. taking antilogs), the final and desired result being the value of pq. This is the process employed in multiplying with the aid of a slide rule.

In calculating the response of a system to an input, we are faced with the problem of integrating the differential equation of the system so as to obtain both the Complementary Function and the Particular Integral, the two portions into which the total response may conveniently be divided**. This integration can be carried out in several different ways. One such way is through the use of the Convolution Integral (sometimes also known as the Superposition Integral or Duhamel's Integral). Providing we can make a start by determining the response of the system to an impulse delivered at time t = 0, this so called impulse response is reversed in time and then multiplied by the input signal and integrated between suitable limits. Readers who are competent at integration, would no doubt manage to solve such problems and obtain correct answers.

However it is usually much easier to take the Laplace transform of the input and to determine the transfer function of the system. The latter may be calculated from the rules of electric circuit theory or by more generalized rules quite simply and directly from the differential equation itself. The former may be obtained from the voluminous tabulations of Laplace transforms which are readily available, provided we know the algebraic form of the input signal as a function of time. Having obtained these two quantities we can then apply the simple operation of multiplication instead of the more difficult operation of convolution integration. Using the initials L.T. to indicate Laplace transform and d.e. for differential equation we have

$$\begin{pmatrix}\text{L.T. of}\\ \text{response}\end{pmatrix} = \begin{pmatrix}\text{Transfer Function}\\ \text{corresponding to}\\ \text{system d.e.}\end{pmatrix} \times \begin{pmatrix}\text{L.T. of}\\ \text{input}\\ \text{signal}\end{pmatrix} \qquad \text{....(10-6)}$$

* A *damped sinusoid* in the variable t (i.e. time) is a sinusoid whose amplitude decreases exponentially with time. It is described by $A\,e^{\sigma t}\cos\omega t$. By a fairly simple re-arrangement it may also be described by the expression $A\,e^{-st}$ where $s = (\sigma + j\omega)$ is called the complex frequency of the signal. Compare this with the artifice of representing $\cos\omega t$ by $e^{j\omega t}$. For a damped sinusoid σ must be negative. Positive values of o will be interpreted as exponential growth of the signal.

** Reminder: the C.F. is the transient response; the P.I. is the steady state response. The C.F. results from solving the differential equation excluding the forcing function but including the initial conditions. The P.I. results from solving the differential equation including the forcing function but virtually excluding the initial conditions. Some degree of interchangeability between forcing function and initial conditions may be exercised arbitrarily in the way in which the problem is stated. See Chapter 9, section 9.6

The above equation is written in terms of the complex variable $s = \sigma + j\omega$. The significance of σ is that it represents damping of a sinusoidal variation occurring at angular frequency ω (see the previous footnote).

The transformation has taken algebra over into *"the complex frequency domain"* or more simply *"the frequency domain"*.

The final operation (analogous with taking antilogs) is then to get ourselves back into the *"time-domain"* (i.e. to recover the response or output signal *as a function of time*) and this may be done by applying an Inverse Laplace Transformation* to the L.H.S. of equation (10-8).

$$\text{Thus} \quad \begin{pmatrix}\text{L.T. of}\\ \text{response}\\ \text{say } F_0(s)\end{pmatrix} \quad \begin{pmatrix}\text{via}\\ \text{Inverse}\\ \text{L.T.}\end{pmatrix} \qquad \text{yields} \qquad \begin{pmatrix}\text{Response asa time}\\ \text{function, say } f_0(t)\end{pmatrix}$$

In addition to referring to the extensive tables of Laplace transforms and inverse transforms, it is necessary for those learning the Laplace transform method to understand and to utilize certain rules so that both transient (C.F.) and steady state (P.I.) portions of the solution may be identified and written down.

Reminder: The rules for using Laplace transform algebra are arranged to give both *transient and steady state* parts of the response.
The rules for using phasor algebra are concerned with yielding *only the steady state* part of the response.

Another pair of names which correspond to these divisions are free and forced response

C.F.	–	transient	–	free response
P.I.	–	steady state	–	forced response

The term steady state is based on the idea that after a suitable length of time the transient part will have died away to insignificance and only the steady state part will remain. During the early part of the response however both transient *and* steady state responses will exist and must be added together, with due regard for sign, to obtain the total response.

The Assembly of Blocks Into Block Diagrams

We are now in a position to consider how we may build up systems by assembling single blocks in such a way that the output of one block forms the input of the next

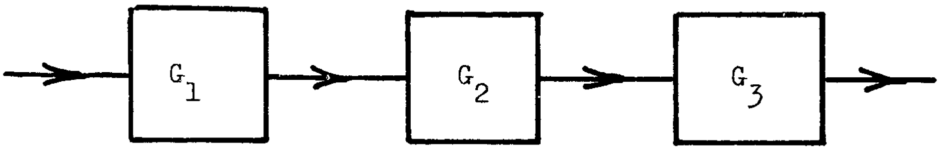


Fig. 10-3

An essential feature (or rule with which we must conform) is that each block is a closed sub-system whose input and output are its *ONLY*** channels for communication or exchange of energy with the remaining parts of the system.

A simplified description of the system represented in Fig. 10-3 is possible and is shown in Fig. 10-4.

* The functions of time, e.g. $f_0(t)$, are sometimes called originals. The corresponding transforms which will be functions of complex frequency, e.g. $F_0(s)$, are sometimes called images. The notation is commonly to employ lower case letters for time functions and the corresponding capital letter for transformed functions. Readers who are not aware of the lower case and capital letter symbols of the Greek alphabet should look up (and learn) such a list in the library.

** Multiple inputs or outputs may be allowed if we can write the differential equations (of each subsystem) in matrix form so that vector arrays may be used for the group of independent (i.e. input) variables and for the group of dependent (i.e. output) variables). The condition then becomes that these two *groups* should form the only channels or connections to the remaining parts of the system.

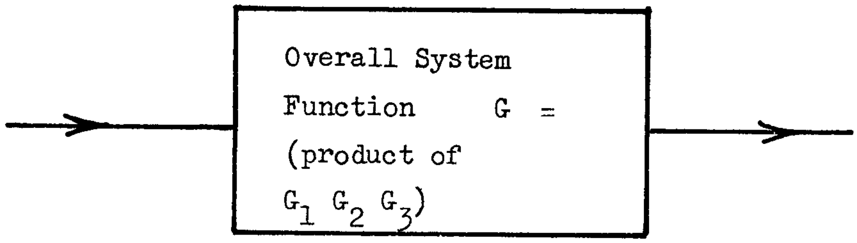


Fig. 10-4

The above simple composition (i.e. putting together) is possible because of the simple multiplicative property once the system function has been written as a function of the complex frequency s, that is as a transfer function. The inputs and outputs in order to be compatible with this simple multiplicative process must also be written as functions of the complex frequency s, that is as Laplace transforms of the actual (i.e. time-domain) signals.

Two* other formal symbols are needed to establish and complete our idea of a block diagram as a formal pictorial representation of a group of differential equations. *The first is the adder or summer.* This is a block with two inputs and one output. See Fig. 10-5. The output signal is the sum of the two input signals.

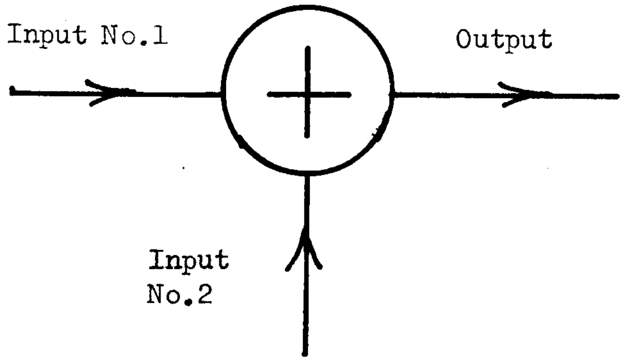


Fig. 10-5

A square block may be used in place of a circle. By reversing the polarity (i.e. sign) of one input, the difference of two quantities may be represented.

Before describing the second symbol, consider how we sometimes go about solving a set of simultaneous equations. One process** widely used is to solve for a certain variable (let us call it x) in equation No. 1 and then to substitute this solution (i.e. result or *output)* into equation No. 2. The position *before* carrying out the substitution may be represented by Fig. 10-6.

The *position* after carrying out the substitution may be represented by Fig. 10-7, which shows first the process of and then the result of the substitution. The result is that the variable x has been eliminated.

* *Other* symbols occur in the literature. Often these will be needed when drawing pictorial representations of mathematical models set up to deal with *non-linear* systems. Examples of this would be a block which multiplies two input variables or a block which produces a transcendental function or a power series of the input variable. However it is not necessary to consider such non-linear models at this stage.

** The aim of the process is of course to *eliminate* one by one the variables in question, ending up with one equation so constituted that it can be solved without ambiguity for a *single remaining dependent variable.*

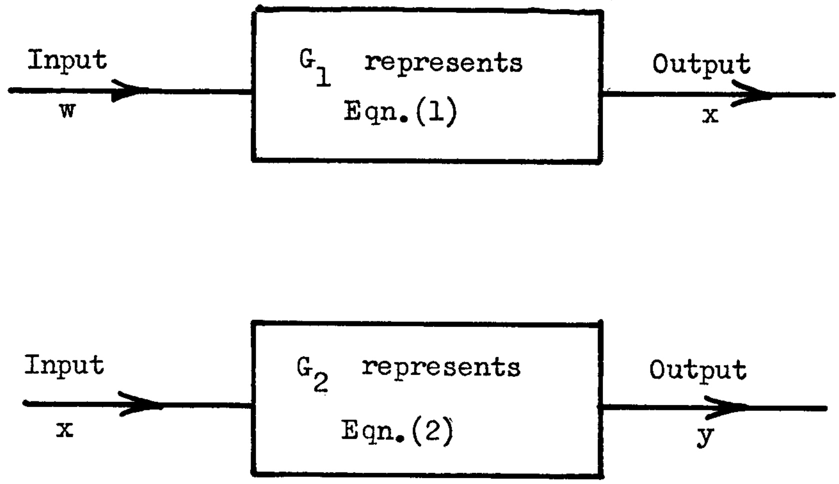


Fig. 10-6

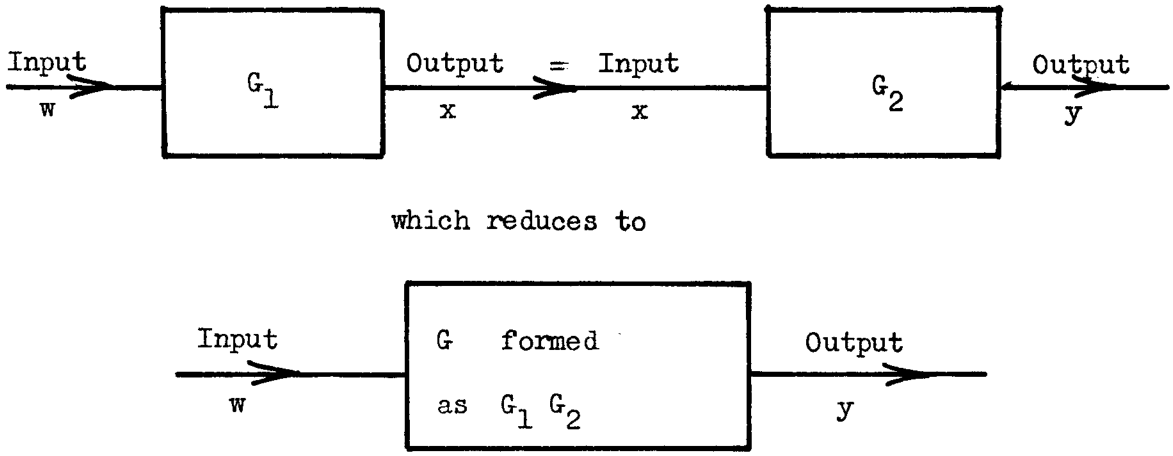


Fig. 10-7

The fact that block diagrams can be utilized in this way to solve quite complicated sets of simultaneous differential equations, depends of course on the formalism which demands that both signals and systems are written in a certain way (as described earlier in this chapter) so that blocks connected in series* may have their system functions multiplied together.

Having thus set the stage, the *second formal symbol* which is simply *the directed line,* can now be introduced. A line connecting two (or more) points simply indicates that the same variable quantity appears at these two (or more) points. The direction of the arrowhead indicates whether the variable in question is independent (arrowhead entering a block) or dependent (arrowhead leaving a block). See Fig. 10-8.

One other possible connection remains. In addition to connecting blocks (subsystems) in series (concatenation) it should be possible to connect blocks in parallel. This is shown in Fig. 10-9.

A little thought will show that two blocks should not be put in parallel without the adder**. Doing so would violate the principles which we have set up for drawing pictorial representations of sets of equations.

Reminder: The representation as a block diagram of a number of sub-systems connected to form a larger system is subject to a few formal rules. These rules although very simple must *NOT* be disobeyed.

* Sometimes the phrase "in concatenation" is used to describe this manner of connection.

**or a subtractor.

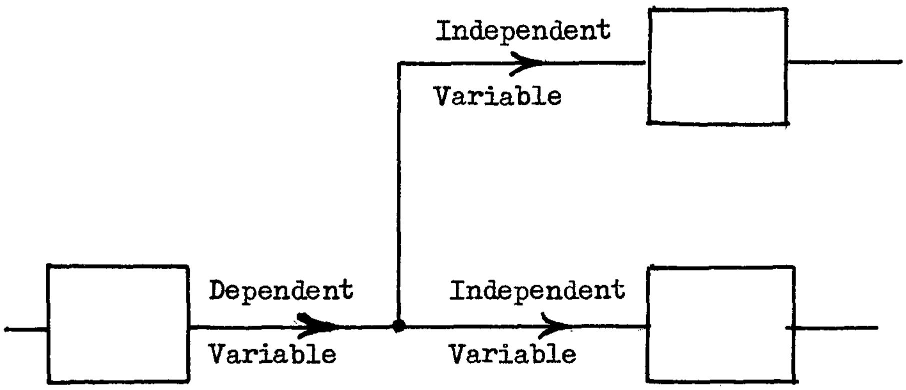
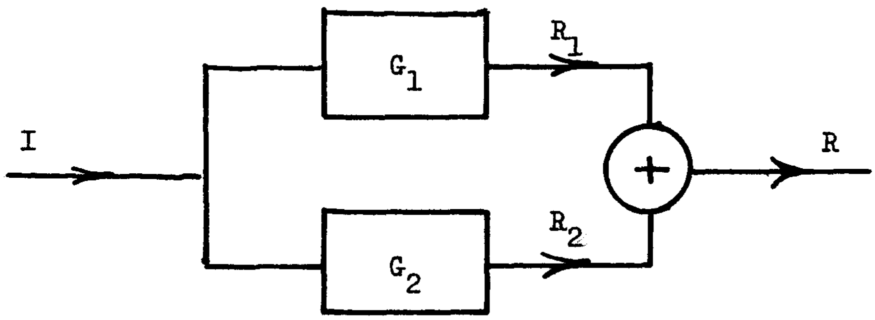


Fig. 10-8



which reduces to

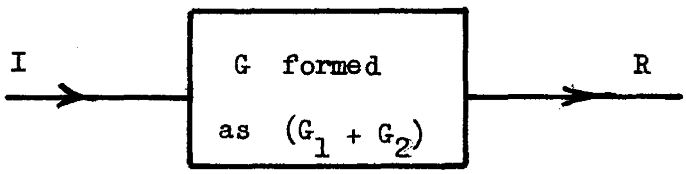


Fig. 10-9

The block diagram representation is also very suitable for describing in pictorial form the mathematical model of a feedback control system. Such diagrams are introduced in Chapter 16. It should be noted that a block may be represented by either a rectangle or a circle. Whichever symbol is used, it is necessary to know of what goes on within the block (i.e. within the subsystem or within the sub-set of equations) either in terms of a system function* or in terms of an equation (or set of equations) ≠. The significance of the circle or rectangle is simply that the subsystem is *closed* and is *connected with* the *remainder* of the system *only by the designated variables* (i.e. connecting lines as designated or shown *on the diagram*). The equations themselves will have been set up by applying suitable conservation laws.

There is one particular configuration of block diagram to which great importance must be attached. This is shown in Fig. 10-10 and it will be noticed that the chain of blocks has been turned back on itself to join up with the input. This configuration is known as a feedback system or feedback control system.

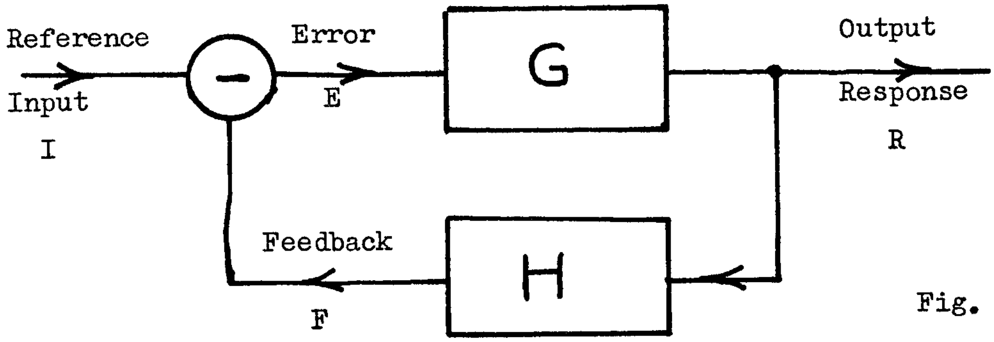


Fig. 10-10

Working with transformed signals we can write down the equations

$$R = G\,E \qquad \text{(from Block G)}$$

$$F = H\,R \qquad \text{(from Block H)}$$

$$E = I - F \qquad \text{(from the subtractor)}$$

Eliminating E and F from these equations

$$R\left(H + \frac{1}{G}\right) = I$$

or

$$\frac{R}{I} = \frac{1}{H} \quad \frac{1}{\left(1 + \frac{1}{GH}\right)}$$

If the product G H (called the loop gain) is large then

$$\text{Approximately } \frac{R}{I} \approx \frac{1}{H}$$

This means that the ratio of output to input can be controlled very accurately to be $\frac{1}{H}$. This quantity can be made to be very stable using low power components of high stability. In other words, the response of a feedback control system can be made practically independent of the forward path (G) which may include a power amplifier of dubious stability.

Often the feedback path (H) is arranged so that H = 1. In that case $\frac{R}{I} \approx 1$ and can be made as close to 1 as we please provided we can arrange a very high gain in the forward path (G). This arrangement is the basis of many automatic regulators.

The arrangement with both G and H represents a widely used system known as negative feedback which can stabilize the gain and decrease the distortion level in an amplifier or modulator. For example if H = 0.05 and if G were made large in value, then the ratio $\frac{R}{I} \approx 20$ and this is the effective gain of the amplifier with feedback. We trade gain for improved stability, lower distortion and improved bandwidth.

* This could be either a transfer function or a phasor response function. The latter is sometimes termed the frequency response.
≠ From which may be derived a system function for the block in question.

10.4 Linear Circuit Theory

Another set of rules intimately associated with a particular kind of pictorial representation has been built up as an aid to solving the integro-differential* equations which describe the electromagnetic behaviour of various electrical devices or networks of such devices. The techniques of circuit theory depend (or may be shown to depend, if an appropriate viewpoint is taken) on the conservation and interaction of stored electrical energy (i.e. stored in an electric field), stored magnetic energy (i.e. stored in a magnetic field), and dissipated energy (i.e. due to imperfect electric conductivity which involves conversion to thermal energy).

Arrangements to cope with conversion of energy to and from chemical, mechanical etc. forms of energy can be set up so that electric circuits may include such devices as batteries, electric motors, generators etc. Because of its dependence on the law of conservation of energy, analogies based on electric circuit theory may be set up for mechanical systems, thermal systems, etc. Indeed wherever*' conservation of energy is the controlling conservation law in a system, we may usefully set up a circuit theory type of representation. Thus we may draw mechanical circuits, thermal circuits, acoustic circuits etc. and use them to simplify the solution of many such problems.

Consider Fig. 10-11. This shows four electrical components with their terminals connected together by lengths of electrically conducting wire.

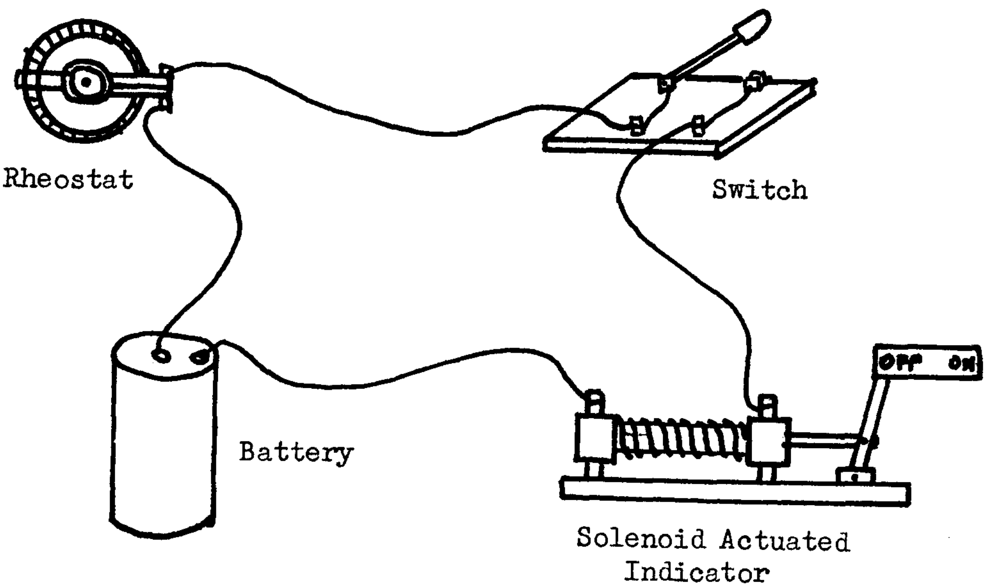


Fig. 10-11

The description of this circuit has been redrawn in Fig. 10-12 as a wiring diagram which incidentally also suggests a neater layout for running the wires between components. Conventional symbols are used for each component.

**i. e. equations in which both differentials and integrals occur.*

*'Subject of course to similar conditions which are imposed on electric circuits. These include for example not trying to operate the system at relativistic velocities' nor at accelerations such that the rigidity of individual mechanical connections is impaired. The limitation to non-relativistic velocities also applies to electric circuits of course if they are in motion while over-large acceleration of electric charges will result in the *radiation* of electromagnetic energy, as from a radio transmitting aerial.

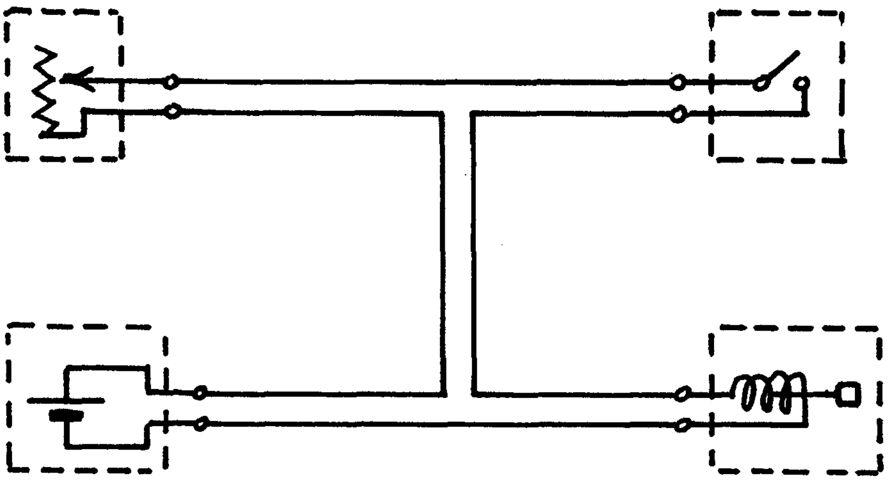


Fig. 10-12

In Fig.10–13 the arrangement is redrawn as a conventional circuit diagram in which there is no attempt to display the relative positions of the components nor the layout (i.e. "run") of the wiring.

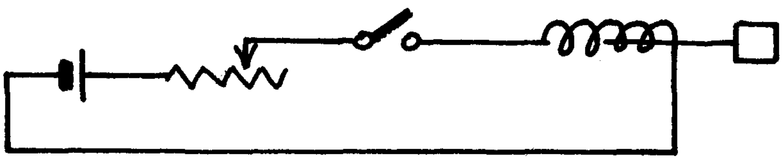


Fig. 10-13

The last diagram is well adapted to explain the electrical operation of the small "system" which has been made up of the four components. It is also well adapted to display one of the two conservation laws on which electric circuit theory is based. These are the two laws known by the name of Kirchoff – one dealing with voltage, the other with current. As you already are aware the Kirchoff voltage law states that the sum of voltages with due regard to sign around any closed circuit is zero; the Kirchoff current law states that the sum of currents flowing into and out of a current junction is zero*.

There is no actual "junction" in Fig. 10-12, the same current flows through each of the four components; since the components are connected in *series* the Kirchoff voltage law which is evaluated by going around a closed loop is applicable; the application of the Kirchoff current law in this particular case yields only trivial relationships. When there are *parallel* or *shunt* paths in a circuit diagram, it would be necessary to apply the Kirchoff current law in building up our set of equations (i.e. our mathematical model).

Historically the idea of a *circuit diagram* as a mathematical model preceded by many decades the idea of a block diagram as a mathematical model. This is not surprising since the circuit diagram for electric circuits follows so readily from the wiring diagram showing actual connections between electrical components of a system. The viewpoint that has just been put, may be summed up in a general way which is applicable not only to electric circuits but also to mechanical, thermal and other networks. Instead of speaking about current and voltage (which are applicable in electric circuits only) we make use of the terms "through-variable" and "across-variable". A tabulation of the quantities which are "through-" and "across-variables" for a few kinds of circuits is shown in Table 10-2.

* Flow out of a junction is reckoned to be of opposite sign to flow into the junction.

TABLE 10-2

KIND OF CIRCUIT	THE THROUGH-VARIABLE IS	THE ACROSS-VARIABLE IS
Electric	Current	Voltage
Mechanical	Force	Velocity
Thermal	Heat Flow	Temperature Difference
Magnetic	Flux	M.M.F.
Hydraulic	Velocity	Pressure Difference
Acoustic	Acoustic Velocity	Acoustic Pressure
Acoustic (inverse)	Acoustic Pressure	Acoustic Velocity

NOTE: In certain text books inverse or so-called dual analogies to the ones listed above may be encountered In such cases not only are the variables interchanged, but the circuit parameters must be re-specified, the circuit configuration looks totally different , and the actual terminals and connections have to be thought of in quite different ways. An example is the so-called classical mechanical circuit analogue where force corresponds to voltage. The mechanical circuit analogue shown in the table above is more useful and is sometimes called the Firestone analogy.

The essence of linear circuit theory lies in the specification of three *passive* circuit components. These would be capacitance, resistance, inductance (C, R and L) for an electric circuit; mass, damping coefficient, spring-constant (M, B and K) for a mechanical circuit. Each of these three is the coefficient* of a particular term in the integro-differential equation modelling the electric or mechanical system.

Taking these in turn,

C or M refer to the differential term in the model

R or B refer to the undifferentiated term in the model

L or K refer to the integral term in the model.

This is illustrated in Table 10-3.

TABLE 10-3

	PARAMETER	DEFINITION	DEFINITION	PARAMETER	
ELECTRIC	C	$i = C\frac{de}{dt}$	$f = M\frac{dv}{dt}$	M	MECHANICAL
	R	$i = \frac{1}{R}e$	$f = Bv$	B	
	L	$i = \frac{1}{L}\int e\ dt$	$f = K\int v\ dt$	K	

In looking at the above table we see that four of the parameters are directly the coefficients concerned. Two others form the coefficient by taking $\frac{1}{R}$ and $\frac{1}{L}$.

* In some cases, one over the coefficient concerned.

The action of ideal components (C, M, L, K) is to store energy. Thus energy may flow either into the component from the circuit or vice versa. The action of the other two ideal components (R, B) is to dissipate energy into heat, whence it is no longer available to flow back into the circuit.

In addition to the passive components, a number of *active* components may be specified. An active component in a circuit corresponds to a forcing function (i.e. a driving source) in the mathematical model. Active devices such as transistors or vacuum tubes may be modelled, in the circuit form* by making use of controlled sources; these are forcing functions which are themselves dependent on some other variable in the system.

The controlled source is also useful for modelling a transducer or transformer. The latter connects two separate sections of an electric circuit by a shared magnetic coupling. A transducer will actually transform energy (usually in *both* directions) from one form into another.

Passive components and ideal sources will each have *two terminals* and *one pair of variables* (a through-variable and an across-variable). See Fig. 10-14.

Transformers and transducers will usually have *four terminals* and *two pairs of variables.* See Fig. 10-15.

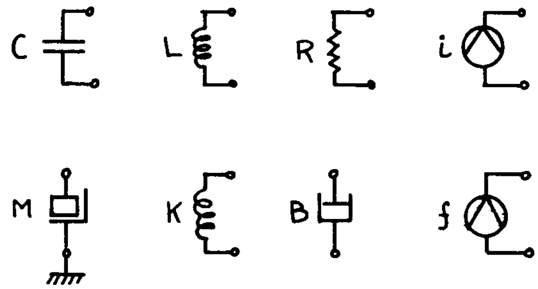
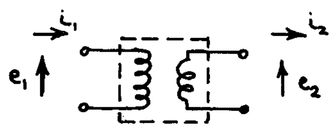


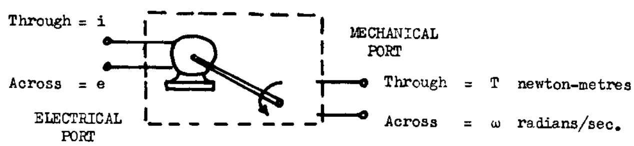
Fig. 10-14



An ideal transformer where (neglecting losses)

$e_1 i_1 = e_2 i_2$ (conservation of energy)

and $\frac{e_2}{e_1}$ depends on turns ratio



An ideal transducer comprising an electric motor where (neglecting losses)

$ei = \omega T$ (conservation of energy)

and the ratios ω/e or T/i depend on the magnetic construction and the windings.

Fig. 10-15

* This procedure is sometimes described as using "an equivalent circuit" for the actual device. A particular device could thus be modelled by a variety of equivalent circuits, depending on what sort of problem we are interested in solving, or in what sort of signal we wish to apply to the system. See reference 10-6-1.

An electric motor where (neglecting losses)

$$ei = \omega T \qquad \text{(conservation of energy)}$$

and the ratios ω/e or T/i depends on the magnetic construction and the windings.

To complete this brief section on linear circuit theory let us note that a differential equation of any order can always be manipulated into a set of simultaneous differential equations of lower order (e.g. first or second order), either by direct algebraic manipulation or by introducing subsidiary variables.

Correspondingly a circuit, by building up L-C-R meshes or loops (each of which is second order) or by using L-R or C-R meshes or loops (each of which is first order), can thus represent or model the dynamics of any lumped system no matter how complicated its interconnections.

We will go on in the next chapter to say something about distributed systems i.e. those systems in which it is not permissible to "lump" the energy storage or energy dissipation effects into separable and discrete circuit parameters.

10.5 Highlights of Chapter 10

(i) Methods of analysing linear systems have been developed using transformed equations and variables.

(ii) Two of such methods involve the use of phasor algebra or the Laplace transform. Both these schemes make use of complex quantities.

(iii) The methods are useful because they replace more complicated operations by simple multiplication or division.

(vi) Associated with these methods is the idea of the block diagram as a graphical model of a dynamic system.

(v) Also associated with these methods is the idea of the circuit diagram as a graphical model of a dynamic system.

(vi) A circuit diagram may be constructed for any dynamic system possessing two (coupled) mechanisms for energy storage and one mechanism for energy dissipation.

(vii) Circuit analogies are usually built up in terms of electric circuits, probably because electric circuit ideas were developed some decades prior to the ideas of mechanical, thermal etc. circuits.

(viii) Either block diagram or circuit diagram models may be solved for transient or steady state responses.

(ix) The transient response is the part represented by the Complimentary Function solution of the differential equation for the system.

(x) The steady state response is the part represented by the Particular Integral solution of the differential equation for the system.

(xi) The total response is obtained by adding together the transient and steady state responses.

10.6 Exercises

1. Two springs are available having spring constants K_1 and K_2 newtons per metre. Prove that when they are connected in parallel they may be replaced by a single spring with constant $(K_1 + K_2)$; or if in series, then by a single spring of constant $\dfrac{1}{{}^1/_{K_1} + {}^1/_{K_2}}$

2. An input signal [30 sin 314t] volts is applied to a black box whose transfer admittance expressed as a phasor quantity is magnitude 0.02 mhos, phase shift +0.5 radians. Determine the output current of the black box, in magnitude and phase. Write down the algebraic expression for the output signal.

   [HINT: Current (amps) = voltage x admittance (mhos)]

3. The input signal to a block box is $f(t) = 2e^{-4t}$ amperes. The transfer function of the black box is

$$\frac{10}{(s+2)} \quad \text{ohms.}$$

Determine the Laplace transform of the output signal. Write down the algebraic expression for output current of the black box expressed as a function of time.
(Assume that by reference to a table of Laplace transforms you have discovered that

(a) the transform of e^{-at} is $\frac{1}{(s+a)}$

and (b) the inverse transform of $\frac{1}{(s+a)}$ is e^{-at})

Hint: If you encounter $\frac{1}{(s+a)(s+b)}$ expand it into partial fractions $\frac{A}{(s+a)} + \frac{B}{(s+b)}$. The inverse transforms of each term in turn may then be written down.

4. Repeat Exercise 9.10.3

10.7 Suggested Reading Material

1. W.A. Lynch and J.G. Truxal, "Introductory System Analysis" (McGraw-Hill – 1961) (preface; chapters 1 and 2; highlights out of chapters 3 and 4).

2. Samuel Seely, "Electromechanical Energy Conversion" (McGraw-Hill 1962) (first chapter only).

Chapter 11

DISTRIBUTED SYSTEMS AND FIELDS

"…. whose virtues are the definitions of the
analytic mind, …. "

("The People", W.B. Yeats, 1919)

Chapter 11

DISTRIBUTED SYSTEMS AND FIELDS

11.1 Introduction

In this chapter it will be pointed out that the notions of systems theory may be extended to what are generally known as distributed systems. In particular we wish to emphasize the point that circuit theory can be used to tackle a large number of problems in distributed systems. However when circuit theory fails (and even in some other special cases) we may have to call on the more powerful (and often more difficult) methods of field theory. A few of these methods will be named.

By a *distributed system* is meant a system in which the *properties of* the system are distributed in such a way that they *cannot be lumped* into separate packages with each package representing the effect of a particular kind of energy storage (or energy dissipation).

If instead of considering the energy within the system (i.e. in order to classify the difference between a distributed and a lumped system) we had focussed our attention on the way in which a suddenly applied forcing function (e.g. in an electrical system, the closing of a switch)affects the various parts of the system we would have come up with a different answer. This answer would have run something like this – in a lumped system there is some sort of instantaneous response (although perhaps the rate of rise is limited) even in the most remote part of the network; in a distributed system we must allow in addition a certain time* for the initial disturbance to spread through the system; at a remote point *nothing at all* happens until after the propagation time has elapsed. In a lumped system the propagation time is reckoned as zero.

If instead of considering energy interchange, or ideas or propagation time we had focussed our attention on the mathematical model for the two kinds of systems we would have seen a very clear and sharp dividing line.

Lumped systems are described** by *ordinary differential equations.*

Distributed systems are described** by *partial differential equations.*

Some of the readers of this book may not yet have encountered partial differentials. These operators are used to describe the behaviour of functions which depend on more than one independent variable. In the case of distributed systems we have to consider not only variation with time but also variation with spatial position. Partial differentials† are written using one form of the Greek small delta i.e. $\frac{\partial}{\partial t}$ $\frac{\partial^2}{\partial t^2}$, etc.

Suppose we were concerned with variations in time and in one space co-ordinate, say x. We would be concerned with partial differentials with respect to t and partial differentials with respect to x. The notation and the meaning would be

$\frac{\delta}{\delta t}$ meaning $\left.\frac{\delta}{\delta t}\right|_{x\,=\,\text{const.}}$ i.e. rate of time-variation when x is held constant

and $\frac{\delta}{\delta x}$ meaning $\left|\frac{\delta}{\delta x}\right._{t\,=\,\text{const.}}$ i.e. rate of variation in x direction when t is held constant.

Such problems are naturally somewhat more difficult to solve. Since many of the readers of this book have not yet dealt with the mathematical ideas and manipulations, we will proceed at first by thinking about some electric circuit theory problems which are "distributed" in one dimension only. Then it will be pointed out that there exist many distributed systems which are non-electrical in character but which may be modelled by the same group of partial differential equations.

A brief reference will be at the end of the chapter to field theory problems where variations may occur in either two or three spatial dimensions.

* the so-called propagation time.

** or by sets of simultaneous such equations in more complicated cases.

† whereas ordinary differentials are written with our normal letter "dee" such as $\frac{d}{dt}$ etc.

11.2 The Uniform Transmission Line

One of the most common examples of the uniform electric transmission line is two long parallel circular metallic wires which have therefore a *uniform* spacing and diameter.

Such a system may store energy in either an electric field or in a magnetic field* and may dissipate energy due to the imperfect electrical conductivity and due also to the imperfect insulation afforded by the medium in which the wires are immersed. This specification leads us to the two partial differential equations

$$\frac{\partial v}{\partial x} = -Ri - L\frac{\partial i}{\partial t} \qquad \ldots.(11.1)$$

$$\frac{\partial i}{\partial x} = -Gv - C\frac{\partial v}{\partial t} \qquad \ldots.(11.2)$$

where the coefficients represent the properties per unit length suggested in Fig. 11-1, which is a circuit model of an infinitesimal length Δx of the line. Note that the coefficient G represents a conductance per unit length (mhos per metre) rather than a resistance.

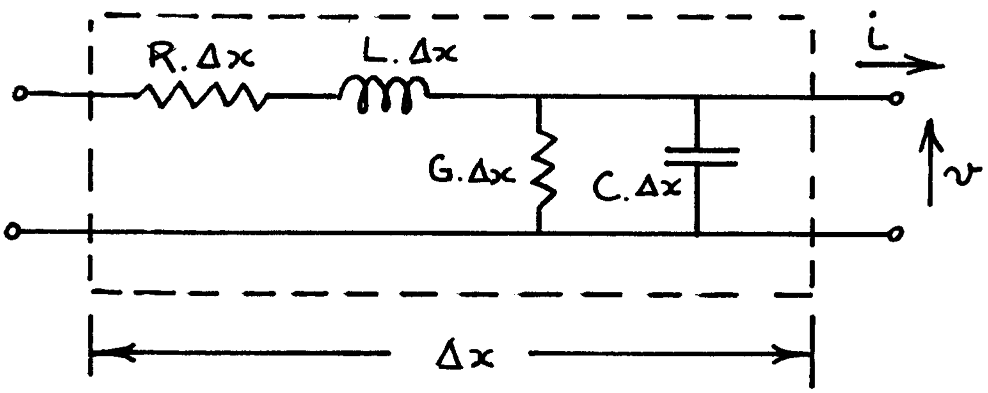


Fig. 11-1

11.2.1 The Telegrapher's Equation

The above pair of equations may be manipulated† into

$$\frac{\partial^2 v}{\partial x^2} = RGv + (RC + LG)\frac{\partial v}{\partial t} + LC\frac{\partial^2 v}{\partial t^2} \qquad \ldots.(11.3)$$

or into an identical equation in which v is replaced by i.

* After reading more advanced texts it will be realized that the separation of energy storage into these two components is possible only in certain cases where we have uniform simple structures and/or particularly simple driving sources, (forcing functions).

† by choosing either to eliminate i or to eliminate v in the pair (11-1), (11-2).

In the form (11-3) the telegraphers equation is a fairly general form of a class of equation known as wave equations. Space variations are written in one group, time variations in another group of terms. Also since the examples being considered are linear, all the coefficients in the equation will be constant quantities.

Interestingly the equation was named the telegrapher's equation since it was first formulated and solved by Lord Kelvin about a century ago in order to explain the limitations of signalling speed encountered on the early trans-Atlantic telegraph cables at that time.

The equation is however of considerable importance elsewhere – it may be used as the mathematical model for many situations including

(i) transmission of electric signals along a uniform line

(ii) mechanical waves on stretched strings

(iii) plane acoustic waves (gases, liquids solids)

(iv) seismic waves (of other types besides the acoustic compression-rarefaction wave in a solid medium)

(v) heat conduction in solids

(vi) mass diffusion (this applies to gases, liquids and solids)

(vii) charge diffusion in semi-conductors

(viii) certain cases of the Schrodinger equation in quantum physics.

11.2.2 Methods Available for Solution

Electrical engineers make use of a number of methods for solving the partial differential equations which they encounter and which may be classed as derivative from the telegrapher's equation. These methods include

(i) the use of *phasor algebra,* with the addition of a graphical aid known as the Smith chart. The technique can cope with uniform lines to which have been connected lumped components. The method is excellent for sinusoidal waveforms in loss-less or slightly lossy lines (i.e. where R and G are zero or small in value) and provides a *steady state sinusoidal solution.*

(ii) the use of the Laplace transform. The technique provides exact solutions *(both transient and steady state portions)* for the lossless case (i.e. where $R = 0$, $G = 0$) and for the particular case known as the distortionless* line (i.e. where $RC = LG$). The method *fails in the general case* because irrational terms appear in the transfer function of the system. Another particular case of great importance is where $L = 0$, resulting in equations describing a number of diffusion processes; a possible solution in this case is the queerly named error function or erf x. ≠

(iii) the use of numerical integration or relaxation techniques. Both these schemes are well adapted for use with a digital computer although relaxation methods may require excessive storage space in the computer if we concern ourselves with more than a few dozen points in the problem region.

(iv) the use of certain other mathematical procedures such as perturbation, the stationary phase principle, contour integration schemes other than Laplace transforms, etc. Procedures such as these are not always specifically included in engineering undergraduate courses†

* In the general case where $RC \neq LG$ all waveforms (except a few special cases such as sinusoids) will become distorted during transmission through a distributed system. Typically, rapidly changing portions of a waveform become spread out or dispersed in time (i.e. slowed down).

≠ A brief historical account of the development of the method nowadays known as the Laplace transform method is given by Carslaw and Jaeger p297-8. They point out for example (p50-52) that there are at least seven types of solution for the simple diffusion equation $\frac{\partial v}{\partial x^2} - \frac{1}{K}\frac{\partial v}{\partial t} = 0$. See reference at the end of this chapter.

† As the complexity and size of engineering systems increase, the value of other mathematical procedures will be recognized. Only occasionally will these procedures *supplant* previously used techniques for mathematical modelling. The scope of engineering will thus seem continuously to increase. One should not be discouraged by this. Most professional people find that it is not too hard to learn to live with this situation, i.e. continuous and rapid growth of concepts and ideas as well as hardware. Indeed it can be a strong source of intellectual stimulation and gratification.

11.3 An Historical Viewpoint

Sometimes ideas and concepts permitting analysis and insight precede the construction and use of the hardware. However it is a fairly commonly held opinion that the invention often comes first and will be followed by appropriate analysis and mathematical models. For example may we quote some remarks made on the occasion of opening the fifteenth Brooklyn Polytechnic International Symposium in New York in 1967; a symposium which was devoted to Systems Theory.

Dr. Emmanuel Piave (Vice-President, IBM) writes:–

> "Industry will sort of roll its sleeves up and move ahead. It always does, whether it understands something or not. The market place will determine whether a thing works, whether it is viable or not".

Dr. Ernst Weber, President Brooklyn Polytechnic Institute:

> "System Theory, particularly as it might be extrapolated to population problems, to economic and biological systems, and others, indicates the penetration of quantitative thought into areas to which engineering has already contributed such basic concepts as feedback, information content, realizability,modelling,etc. . . . . . . The broadly ranging concepts and tools of System Theory may serve to accelerate the trends discernible now but still shrouded in early morning mist."

Dr. W.G. Shepherd, Vice President Institute of Electrical and Electronic Engineers, N.Y.:

> "World War II alerted the technological world to the need for systems capability. The initial system efforts of that time were largely ad hoc and fragmented."

These extracts seem to be saying that analysis and insight *follow* construction and use. Let us give historical examples where the reverse has sometimes been true. For this let us trace some developments in electrical transmission lines.

First of all, some dates are listed

About 1830	early attempts at telegraph transmission by people such as Gauss and Weber in Europe.
1844	the Morse telegraph in the U.S.
1851	undersea telegraph cable, England-France
1866	undersea telegraph cable, trans-Atlantic
1876	telephony needing wider bandwidth capabilities in the transmission lines
1886	invention of the polyphase induction motor by Tesla and the *need* (or the possibility?) for transmission of electric power over polyphase transmission lines leading to efficient *central* generating stations supplanting local generation.
1901	Marconi successfully transmits a radio signal across the Atlantic. Transmission lines needed to join the equipment to the aerials.
About 1940	Widespread use of hollow pipe waveguides when microwave sources of power became "commercially" available.

If one looks at these one may see successful telegraph lines antedating Kelvin's mathematical analysis of the telegrapher's equation (about 1866). However one may also see the much greater success of later telegraph lines and telephone lines which had been "loaded" (i.e. with increased inductance) in order to increase the speed and frequency of operation*. On the other hand Marconi's successful construction of radio stations came many years after the prediction (by Clerk Maxwell) and the discovery (by Heinrich Hertz) of the existence of radio waves. Also we may note the theory of guided electromagnetic waves as available many years (early this century) before the application of waveguides in the early microwave radar systems of World War II.

* Readers should see what they can discover in the library about this. Among others look for mention of Kelvin (England), Bell (U.S.A.), Pupin concerning information transmission; look for mention of Tesla (U.S.A.), Siemens (Germany) and Ferranti (England) concerning power transmission.

This seems confusing unless we can understand that the *application* and *development* of modern technology proceeds rather in a leap-frog fashion, with advances in theory, discovery by trial and experiment, improvement by acquaintance and operating experience all contributing their share from time to time (and not necessarily in any fixed order nor in any fixed proportion) to the next developments.* The various phases mentioned here proceed, not necessarily in any *predictable* order but nevertheless each will contribute *understandably* to the later advances.

11.4 Field Theory

In field theory we are concerned with more complicated partial differential equations, which may be written in two or three spatial coordinates (or in one dimension, although we have discussed some special cases of this for Cartesian Coordinates under the heading of the Telegraphers Equation).Table 11-1 below suggests that one way of looking at field theory problems is to classify the equations by the way in which variations with time appear in the equation – i.e. whether terms in $\frac{\partial^2}{\partial t^2}$ or $\frac{\partial}{\partial t}$ appear or not (in more complicated cases combinations of these terms may crop up).

Table 11-1

CLASS OF PROBLEM	WRITTEN IN VARIABLES	NAME OF EQUATION	TYPICAL APPLICATION
STATIC problems	1, 2 or 3 space variables	Laplace and Poisson equations	High voltage insulation.
DIFFUSION problems	1, 2 or 3 space variables and $(\frac{\partial}{\partial t})$	The Diffusion equation	Diffusion of water through an earth dam.
WAVE problems	1, 2 or 3 space variables and $\frac{\partial^2}{\partial t^2}$	The Wave equation (loss–less medium)	Radiation from aerials. Propagation of sound underwater.

The actual equations tend to be rather complicated. For example the simplest one is the Laplace equation. Written in three spatial variables using spherical coordinates (r, θ, ψ), it is

$$r\frac{\partial^2\Phi}{\partial r^2} + 2\frac{\partial\Phi}{\partial r} + \frac{1}{r}\frac{\partial^2\Phi}{\partial\theta^2} + \frac{1}{r\tan\theta}\frac{\partial\Phi}{\partial\theta} + \frac{1}{r\sin^2\theta}\frac{\partial^2\Phi}{\partial\psi^2} = 0$$

* Readers should think again of the analogy presented earlier by one of the other authors of this series, where the process of technical innovation was likened to the process of a child playing with its toys. One will have observed the actions of different children at play. Some occupy themselves with a busy series of random trials; others are more contemplative and make pictures in their minds before acting out the situations they have created in their imaginations; others again will seize on a theme, in which creation by imagination proceeds step by step alternatively with movements of the toys and differing involvements of the child itself with the patterns formed by the boys. Readers should use their imagination too to discern patterns of social behaviour created and carried on by adolescents and adults of their acquaintance

where ϕ is used to indicate a potential. A *shorthand* approach which is more generally applicable (i.e. to a variety of coordinate systems) is provided by a branch of applied mathematics known as vector analysis. Using the symbolism of vector analysis the last rather nasty looking equation reduces to $\Delta^2\phi = 0$ where the Δ^2 is known as the Laplacian operator.

If we are content with sinusoidal signals then the diffusion and wave equations may be reduced to a form known as the Helmholtz equation. Solutions to both the Laplace and the Helmholtz equations are tabulated in handbooks or in publications devoted to more individual problems.

Certain special functions are also tabulated for use in carrying through the solutions. These are associated with patticular coordinate systems and you will certainly be familiar with the first one listed in Table 11-2.

Table 11-2

COORDINATE SYSTEMS	TYPE OF SPECIAL FUNCTION
Rectangular	Trigonometric (i.e. Sin, Cos)
Spherical	Bessel functions
Cylindrical (circular)	Legendre functions
Cylindrical (elliptic)	Mathieu functions
Cylindrical (parabolic)	Weber functions
Conical	Lame functions

With these aids we may analyse quite a number of field problems, provided that the boundaries of the problem are not too complicated in shape. When the problem won't fit in with the above analytical methods (e.g. because the hardware doesn't fit onto one of the available and usable coordinate systems) it may be possible to go to numerical methods or to a method using the complex variable known as conformal mapping (often suitable for two-dimensional problems) or else to experimental methods (suitable for two- and three-dimensional problems). Experimental methods include the use of electrically conducting paper or electrolytically conducting liquids in a "tank".*

11.5 The Maxwell Equations

Using the formalism of vector analysis (i.e. the notation, plus the ideas conveyed by the notation which will be encountered a little later in undergraduate courses) the fundamental actions of practically the whole of electromagnetism may be written down as a set of four simultaneous partial differential equations.

These are the simultaneous equations for electric and magnetic fields, $\overline{E}$ and $\overline{H}$.

Note ρ = charge density, ϵ = permittivity, μ = permeability**

$$\text{Curl } \overline{E} = -\mu\frac{\partial\overline{H}}{\partial t} \qquad \text{div } \overline{E} = \rho/\epsilon$$

$$\text{Curl } \overline{H} = \sigma\overline{E} + \frac{\partial\overline{E}}{\partial t} \qquad \text{div } \overline{H} = 0$$

It was an outstanding feat to have written down in such a compact form, such an enormously generalized set of laws.

* NOTE: One-dimensional field problems may also be solved by a direct integration of the Laplace or Helmholtz equation.

** Curl, divergence, gradient and ∇^2 are the *operators* of the system of algebra known as vector analysis.

Maxwell's equations form the basis of nearly every analytical attack on electromagnetic problems. Most textbooks on electromagnetic theory consider only rigid conductors, either stationary or moving. However texts dealing with flexible and fluid conductors have also been written, an example being reference (11-5) in the suggested further reading.

11.6 Highlights of Chapter 11

(i) In distributed systems, the propagation time or travelling time of a signal is significant.

(ii) Distributed systems are described by partial differential equations.

(iii) The telegrapher's equation for a uniform transmission line is also a model for many other diffusion and wave processes.

(iv) Field theory problems are often expressed by using the operators and symbolism of vector analysis.

(v) Maxwell's equations can represent the behaviour of practically any electromagnetic system.

11.7 Exercises

(1) Derive equation (11-3) from equations (11-1), (11-2).

(2) On page 337 of reference (11.4) a distinction is made between a plane wave travelling in a "good" dielectric and in a good conductor. Try to build up the analogy between these cases and a uniform transmission line. To what ranges of value for R,L,G and C would these two cases correspond?

(3) In this chapter is given a list of dates signifying steps in the successful technical development of electrical transmission lines. Make a similar list with approximate dates to show alternate steps of theory and practice contributing to the development of one or two other modern technologies.

(4) In section 3 of Chapter 11, comment is made about loading telegraph or telephone lines to increase inductance. See what you can discover in the library about this. Among other names look for any mention of Kelvin (England), Bell (U.S.A.), Pupin. See also what you can discover about early power transmission and in this case look for mention of Tesla (U.S.A.), Siemens (Germany), and Ferranti (England).

11.8 Suggested Reading Material

(11-1) R.K. Moore "Wave and Diffusion Analogies"

(McGraw Hill 1964) – Chap. 1, 2, 8, 10.

(11-2) H.S. Carslaw and J.C. Jaeger "Conduction of Heat in Solids"

(Oxford University Press, 1959) – Pages 1-3, 50-52, 297-8.

(11-3) Phillip Magnusson "Transmission Lines and Wave Propagation"

(Allyn and Bacon, 1964) – preface, introduction and some of Chapter 7, particularly the overview on p.160.

(11-4) S. Ramo, Whinnery and van Duzer, "Fields and Waves in Communication Electronics", (Wiley International, 1965) – pages 330-337.

(11-5) J.A. Shercliff "A Textbook of Magnetohydrodynamics"

(Pergamon, 1965) – Chapter 1 and a small portion of Chapter 2.

Chapter 12

HISTORY AND APPLICATIONS OF COMPUTERS

DEAR READER, this notice will serve to inform you that I submit to the public a small machine of my invention, by means of which you alone may, without any effort, perform all the operations of arithmetic, and may be relieved of the work which has often times fatigued your spirit, when you have worked with the counters or with the pen. As for simplicity of movement of the operations, I have so devised it that, although the operations of arithmetic are in a way opposed the one to the other—as addition to subtraction, and multiplication to division—nevertheless they are all performed on this machine by a single unique movement. The facility of this movement of operation is very evident since it is just as easy to move one thousand or ten thousand dials, all at one time, if one desires to make a single dial move, although all accomplish the movement perfectly. The most ignorant find as many advantages as the most experienced. The instrument makes up for ignorance and for lack of practice, and even without any effort of the operator, it makes possible shortcuts by itself, whenever the numbers are set down.

BLAISE PASCAL
On His Calculating Machine, 1649

Chapter 12

HISTORY AND APPLICATIONS OF COMPUTERS

12.1 Analog and digital computers

Automatic computers may be classified broadly as analog or digital. An analog computer works with continuous variables and makes use of the analogy between the values assumed by some physical quantity such as an electrical voltage or current, distance, or shaft rotation. The slide rule is an example of an analog computer. On the other hand the digital machine works with numbers directly although it must of course represent such numbers by a physical quantity. However, the analogy only requires that the variable represented should have restricted discrete values. Fig. 12.1 is intended to show a broad classification of computers. In this book our interest will be confined to digital machines.

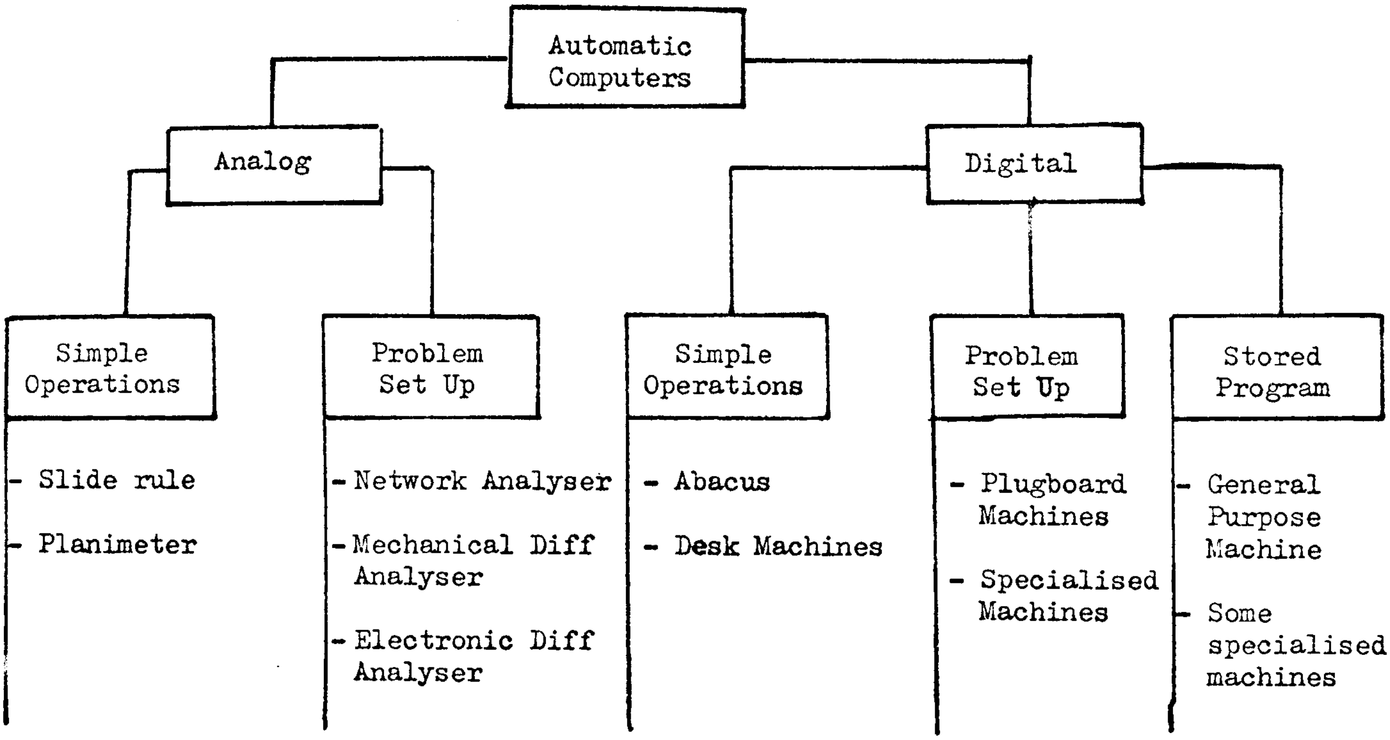


Fig. 12.1 Classification of Computers

12.2 History of digital computers

Early concepts

Mechanical aids to counting and calculating have been known and used for centuries. One of the most ancient, the abacus is still in use today and was known to be in use in China as early as the sixth century (B.C.). However, the first mechanical device capable of addition and subtraction in a digital manner has been generally credited to Pascal (1642). His machines had number wheels the positions of which could be observed through windows in their covers. The stylus-operated pocket adding machines now widely used are descendants of Pascal's machine.

Leibnitz first proposed a machine which could multiply by repeated addition (1671). One was built in 1694 but it was not reliable. The first successful commercial machine capable of all four basic arithmetic operations was the Arithometer of C.X. Thomas (1820). However, such machines were not in common use until the latter part of the nineteenth century.

The Jacquard loom punched card system is also an important forerunner of automatic calculators. This loom was the first successful application of the principles of punched tape and card control and was demonstrated originally between 1725

and 1745 by Bauchen, Falcon and Jacques. It came into widespread use in the decade following 1804. One of the interesting developments in recent years has been in computer control of looms.

The difference engine was another invention of importance – it was a device for automatically calculating mathematical tables of functions whose higher-order differences are constant. This requires a register for each order of difference and a means for adding the contents of each register to those of the next lower order register. The principle is due to Muller (1786) and construction was first attempted by Charles Babbage who between 1812 and 1822 built a small working model.

Charles Babbage was a British mathematician and engineer (at one time, Professor of Mathematics at Cambridge) and is regarded as the most important of the early inventors of computers. It was characteristic of him that having made a small difference engine he immediately proposed the biggest difference engine that might ever be needed. Support for the venture was obtained from the government and the Royal Society. However, the technology of the time was not adequate for the task and the project was still incomplete 10 years later (1833). A commercial version of a smaller machine was later built in Sweden. At this stage Babbage conceived his analytical engine, the first design for a universal automatic calculator, and he spent the remainder of his years working on it (largely with his own money). This machine had all of the conceptual elements of a modern general purpose computer; a memory of 1000 words of 50 digits each (all in counting wheels); an arithmetic mill and a control unit based on sequences of Jacquard punched cards. The important principle of conditional branching of control on the algebraic sign of a designated number was included. This was designed to cause skipping forwards or backwards by a specified number of cards. The arithmetic unit was designed for one addition or subtraction per second or a multiplication in about one minute. Provision was made for access to function tables stored on cards. Babbage prepared thousands of drawings but was only able to complete a few parts of the machine some of which are in the British Museum.

The idea of machine-readable unit record documents is due to Hollerith. He introduced electro-mechanical means of sensing holes in punched cards and apparatus for entering, classifying, distributing and recording data. The development of many types of electromechanical accounting machine was an outcome of his work and his machines were used during the compilation of the 1890 Census reports of the U.S.A.

The Bell Telephone Laboratories developed several relay computers based initially on the work of Stibitz (1938). These machines used paper tape input, program control, branching, self checking, and other features of later electronic machines. However, all operations were carried out with relays.

Transitional machines

Babbage's ideas were realised with the construction of the Automatic Sequence Controlled Calculator, or Harvard Mark I. This was a very large electromechanical calculator (51 ft. x 8 ft.) fuilt between 1939 and 1944 as a joint venture by I.B.M. and Harvard University under the direction of Howard Aiken, Professor of Applied Mathematics. It was a parallel synchronous machine based on a number length of 23 decimal digits plus sign. It could perform addition, subtraction, multiplication, division and reference to tables under the control of an automatic sequence mechanism using a perforated paper "control tape". The fundamental cycle was 3 milliseconds and a typical multiplication required about 3 seconds. It operated successfully for about 15 years and many of the tables in current use were prepared on this machine.

ENIAC was the first electronic computer and was built between 1943 and 1946 at the Moore School of Electrical Engineering of the University of Pennsylvania. ENIAC (Electronic Numerical Integrator and Computer) was not designed as a general purpose machine but was intended to produce firing and ballistic tables for the U.S. Army by solving the corresponding system of differential equations using arbitrary functions. It was most impressive in physical size; it occupied a space of 30 ft. by 50 ft., contained 18,000 vacuum tubes and required 130 kw of power. Addition required 0.2 milliseconds and multiplication 2.8 milliseconds. Although ENIAC was not a general purpose machine it was a major step forward in speed (1000 times) and brought together computers and electronics. A number of limited general purpose electronic machines were produced commercially in the late 1940's. Of these the I.B.M. 604 is probably the best known.

Electronic general purpose computers

The ideas generated during the development of ENIAC were collected together in a report titled "Preliminary Discussion of the Logical Design of an Electronic Computing Instrument" by A.W. Burks, H.H. Goldstine, and J. von Neumann. The report was prepared at the Institute for Advanced Study at Princeton and was widely disseminated by means of a summer session at the University of Pennsylvania in 1946. Very shortly thereafter, electronic computer projects started at a number of institutions throughout the world. At this stage a machine employing the new concepts was already under way at the University of Pennsylvania (EDVAC). The new concept employed was that the program is expressed as a set of instructions encoded and stored in the same store as that used for storage of data and constants. This carries the implication that the machine can operate on its own program as data and thus modify its course of action in a flexible manner.

The first stored-program computer to be placed in operation was EDSAC at Cambridge University (May 1949). This was a binary machine (17 bits) with limited storage (512 numbers) based on mercury delay lines with punched paper tape input and output.

Another early machine was at the University of Manchester – this introduced new principles in storage, the electrostatic storage tube (Williams tube) and magnetic drum, and first made use of index registers for address modification. It also operated in prototype form in 1949.

A number of similar projects were completed elsewhere at about the same time. SEAC at the U.S. Bureau of Standards introduced a number of features in construction and design. Possibly some of the most influential ideas cam from the group at the Institute for Advanced Study at Princeton which gave rise to a family of IAS type machines. These were parallel binary asynchronous machines of which ILLIAC (University of Illinois) and the I.B.M. 701 are decendants. The first commercial machines appeared in 1950-51. Among the first was the ERA 1101 (Engineering Research Associates), UNIVAC and in the U.K. the Ferranti Mark I. A machine built in Australia by C.S.I.R.O. and later termed CSIRAC was, also among the first dozen or so in the world. It was a serial binary machine using mercury delay lines and magnetic drum and operated successfully in 1951.

Modern systems

The hardware realisation of early machines differed markedly from each other – however, the basic concepts of organisation were essentially similar, and have remained so. Thus the subsequent development of computers has been evolutionary rather than revolutionary. The major breakthroughs have been in technology and have led to dramatic improvements in storage, speed, and reliability. The improvements in speed alone have been in excess of 1000 to one. At the same time a great deal has been achieved in reducing the amount of human effort necessary to express problems in a form suitable for computer solution. We will be discussing all of these topics in greater detail in later chapters. Before doing so we will discuss some of the more outstanding uses of computers in modern science, engineering, education, and business.

12.3 Applications of computers

Computers were first used for scientific research and indeed a great deal of the incentive to build machines came from this need. The reason was that several important lines of research were making very little progress because of the requirement for very large amounts of computing. Nuclear physics and astronomy were probably the main areas of need. Astronomers were unable to test new theories on the evolution of stars and to correlate these with observations until computers became available. Nuclear research creates computing problems in almost all of its aspects and nuclear research establishments have the biggest installations in the world. One of the more interesting is in the analysis of tracks in bubble chambers which are produced by charged particles resulting from nuclear collisions. It is not uncommon to require analysis of several million photographs per year each carrying tracks of a number of particles. These tracks have to be analysed to determine the possible existence of a new particle. The structure of crystals can also be determined by patterns resulting from X-rays which are directed at a sample. Analysis and refinement of this data can yield a structure for the crystal, knowledge which is of considerable value to chemists. However, these calculations make very heavy demands on computers, particularly now that small computers are also being used to control the apparatus and collect the data.

Engineers have made a great deal of use of computers. Many engineering projects involve huge expenditure so that design refinement can yield substantial savings. Perhaps the most outstanding example of this is in the aircraft industry where we find some of the biggest computer installations in the world, outside of the nuclear field. A new aircraft design requires the most thorough analysis possible – the price of error can be costly in every way. Computer analysis of frames for buildings, bridges and other structures is also established practice. Study of the motion of fluids is important to many aspects of engineering – aircraft, ships, water flow. These problems can involve a great deal of computing.

Electrical engineers have been using computers for a very wide variety of problems – from studies of electric power distribution to the automatic production of masks for fabrication of integrated circuits. The computer is also becoming an important unit in communication systems – in fact, the new electronic telephone exchanges are based on computer techniques. This is an interesting change since computers relied for their initial development on the components and techniques of the communication industry. Computer control is also becoming established in applications ranging from control of huge power stations to control of aircraft, radio telescopes, industrial processes and space vehicles. As might be expected computers are also used to assemble computers; the production of logic cards, and the connection of their base wiring is now largely done by machines controlled by computers. The role of computers in education is also developing rapidly. The more advanced systems work through personal consoles allowing interaction between machine and student. As more suitable hardware becomes available and the learning process better understood this type of use can be expected to increase very rapidly. The concept and value of personal consoles has already found widespread acceptance and need in many other applications.

Business applications of computers have also found wide acceptance. The traditional role of computers in business is in areas such as payroll, stock and inventory, and general accounting. However, business is now making use of computer methods to improve plant scheduling, to simulate or model business situations, and in other ways designed to optimise current operations and evaluate future plans.

The future of the computer is hard to predict. In less than twenty years it has grown from zero to 100,000 installations and has become an indispensible part of science, engineering and business. At the same time the power of a machine of similar cost has improved by 1,000 to 1. This progress is probably unparalleled in industrial history so that we have now come to accept

computers as a normal part of our lives. There does not seem to be any major reason to suppose that progress will not continue – the main limitation could be in learning how best to use computers, and how to make them even easier to use.

12.4 Exercises

1. Classify the following as being analog or digital: –

   (a) Cash register

   (b) Car speedometer

   (c) Car mileage meter

   (d) Wrist watch

   (e) Car fuel gauge.

2. Suggest some new areas of application of computers which could result from reductions in the cost of machines.

12.5 Suggested reading

1. "Charles Babbage and His Calculating Engines"

   Edited by Philip and Emily Morrison (Dover 61).

2. "Faster Than Thought" edited by B.V. Bowden

   (Pitman).

3. "Preliminary Discussion of the Logical Design of an Electronic Computing Instrument"

   A. W. Barker, H. H. Goldstine and John von Veumann.

4. "The Evolution of Computing Machines and Systems"

   R. Serrell et. al. Proc. IRE May 1962. p. 1039.

Chapter 13

SWITCHING CIRCUITS, NUMBER SYSTEMS AND BASIC COMPUTER ORGANISATION

If you look at automata which have been built by men or which exist in nature you will very frequently notice that their structure is controlled only partly by rigorous requirements and is controlled to a much larger extent by the manner in which they might fail and by the (more or less effective) precautionary measures which have been taken against their failure. And to say that they are precautions against failure is to overstate the case, to use an optimistic terminology which is completely alien to the subject. Rather than precautions against failure, they are arrangements by which it is attempted to achieve a state where at least a majority of all failures will not be lethal. There can be no question of eliminating failures or of completely paralyzing the effects of failures. All we can try to do is to arrange an automaton so that in the vast majority of failures it can continue to operate. These arrangements give palliatives of failures, not cures. Most of the arrangements of artificial and natural automata and the principles involved therein are of this sort.

John von Neumann
Theory of Self-Reproducing Automata, 1966

Chapter 13

SWITCHING CIRCUITS, NUMBER SYSTEMS AND BASIC COMPUTER ORGANISATION

13.1 Switching circuits and logic

Logic

Logic as a subject originated as a branch of philosophy in the fourth century (B.C.). However, progress was relatively slow until 1847 when George Boole and Augustus de Morgan independently developed algebras for symbolically representing and manipulating logical expressions. The same principles of logic are particularly valuable in the analysis and synthesis of switching circuits and in programming. Shannon (1938) was responsible for initial development of methods of symbolic analysis of relay and switching circuits.

A logical variable, sometimes called a Boolean or binary variable is one that can take on only two distinct values. The value assigned thus represents a choice between two alternatives which can be denoted by the symbols 1 and 0. Alternative symbols sometimes used are T (true) and F (false) or X (true) and $\bar{X}$ (not X).

Truth tables

A function of one or more binary variables can be completely specified by a table listing all possible argument values and the corresponding function value for each. Such a table is called a *truth table.*

Table 13.1a

Argument		Function			
x	y	AND	OR	NAND	NOR
0	0	0	0	1	1
0	1	0	1	1	0
1	0	0	1	1	0
1	1	1	1	0	0

Table 13.1a is a truth table which defines the functions AND, OR, NAND, NOR. The AND function is true only when x AND y are true, the OR is true when x OR y is true. The NAND is the negate of AND (NOT AND) and NOR the negate of OR (NOT OR). All are capable of simple realisation in relay or electronic circuits. The set AND, OR, NOT are primitives and complete in the sense that all other functions can be expressed in terms of them. Fig. 13.1a shows some representative circuit realisations for the functions.

A truth table of n input variables has 2^n rows which are usually listed as if they were consecutive binary numbers. On the right of each row a 0 or 1 is entered depending the output value required for the combination of inputs represented by the row. Table 13.1b shows a truth table for 3 input variables. The example is that of binary addition in which there are 3 inputs, x, y, and c (carry in) and two outputs, sum and carry(It is worth noting that for 3 input variables there are 2^8 or 256 possible output functions – in general for n input variables there are 2^n rows and therefore $2^{(2^n)}$ possible truth tables). Sometimes a particular input combination is known to be impossible or an output combination can be allowed as 0 or 1. In such cases a x may be used to assign a "don't care" condition – such conditions can often lead to circuit simplifications.

Boolean algebra

Although any function may be expressed as a truth table it is convenient to have an algebra from which new properties and relations may also be deduced. The set of functions, AND, OR, NOT are convenient primitives and can be used in a way somewhat analogous to multiplication and addition in ordinary algebra. The following points summarise key features. AND is written $\wedge$ and OR is written $\vee$. Alternative symbols are   for AND and + for OR.

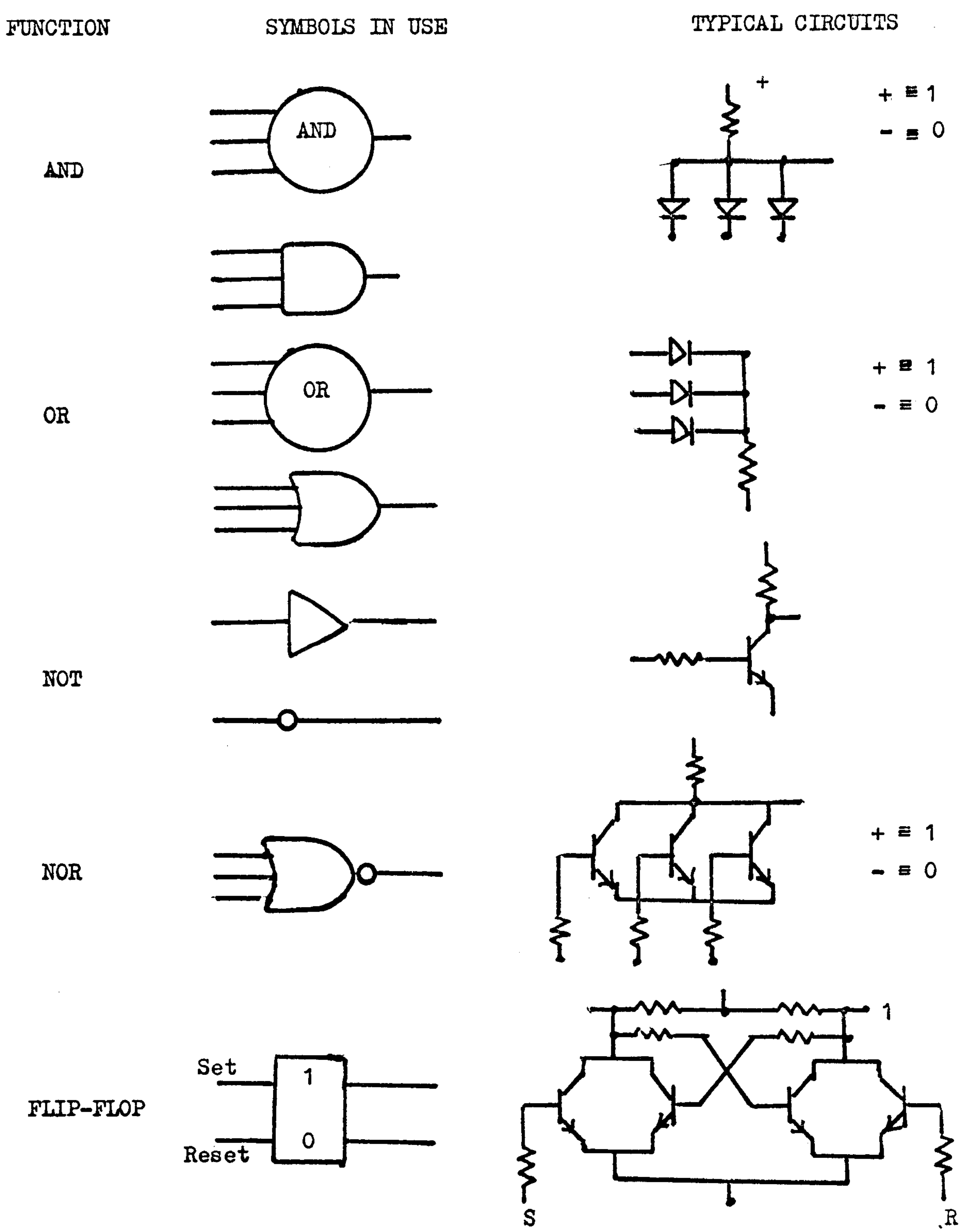


FIG. 13.1a Logic Functions Symbols and Circuits

x	y	c	Sum	Carry
0	0	0	0	0
0	0	1	1	0
0	1	0	1	0
0	1	1	0	1
1	0	0	1	0
1	0	1	0	1
1	1	0	0	1
1	1	1	1	1

Table 13.1 b

(a) The operators AND and OR obey the association law.

$a \wedge b \wedge c \equiv (a \wedge b) \wedge c \equiv a \wedge (b \wedge c)$

$a \vee b \vee c \equiv (a \vee b) \vee c \equiv a \vee (b \vee c)$ ....(13.1 a)

(b) The operators AND and OR also obey the commutation law and distribution laws.

$a \wedge b \equiv b \wedge a$

$a \vee b \equiv b \vee a$

$a \wedge (b \vee c) \equiv (a \wedge b) \vee (a \wedge c)$

$a \vee (b \wedge c) \equiv (a \vee b) \wedge (a \vee c)$ ....(13.1 b)

The following identities may be shown to hold by testing the appropriate truth table.

(c) $a \wedge o \equiv 0$

$a \wedge 1 \equiv a$

$a \vee 0 \equiv a$

$a \vee 1 \equiv 1$

$a \wedge a \equiv a$

$a \vee a \equiv a$

$a \vee \bar{a} \equiv 1$

$a \wedge \bar{a} \equiv 0$

$a = \bar{\bar{a}}$ ....(13.1 c)

(d) $a \vee (a \wedge b) \equiv a$

$a \vee (\bar{a} \wedge b) \equiv a \vee b$

$a \wedge (a \vee b) \equiv a$ ....(13.1 d)

(e) $\overline{a \wedge b} \equiv \bar{a} \vee \bar{b}$

$a \wedge b \equiv \overline{\bar{a} \vee \bar{b}}$

$\overline{a \vee b} \equiv \bar{a} \wedge \bar{b}$

$a \vee b \equiv \overline{\bar{a} \wedge \bar{b}}$ ....(13.1 e)

Equations (e) are forms of De Morgan's law and are especially useful in analysis of logic circuits. These laws are embodied in a number of methods of simplification which are outside the scope of this course. Probably the two most useful methods are the Karnaugh map and the Quine-McCluskey simplification algorithm.

Any truth table may be alternatively written as an algebraic expression (and vice versa). There are two methods yielding different but equivalent algebraic expressions – the first yields an AND-OR form of circuit and the second an OR-AND form.

(a) For each case where the output function is 1 provided an AND gate to detect the state of the input variables. Combine the AND gates in an OR. Fig. 13-1 b shows the result obtained from table 13.1 b.

(b) For each case where the output function is 0 provide an OR gate for the complement of the corresponding input variables. Combine the OR outputs in an AND circuit.

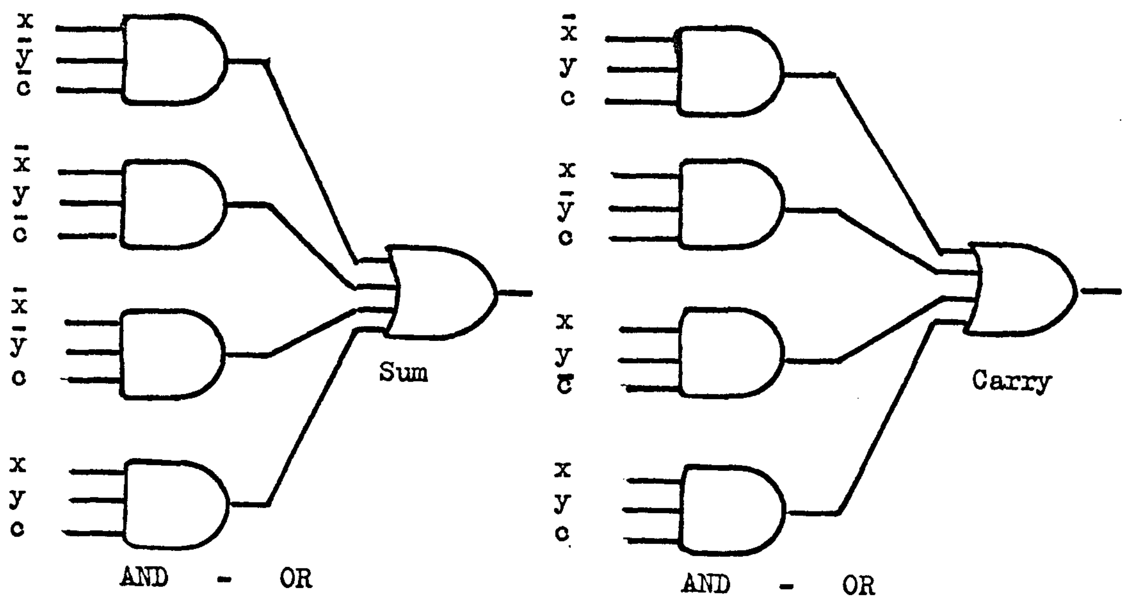


FIG. 13.1b Binary Addition

Before proposing any schemes for arithmetic operations we need to review some of the principles of number representation.

13.2 Number representations and arithmetic operations

Numerical or other types of information are usually represented internally in computer systems by finite strings of two-valued signals (binary digits or bits) denoted by 0 or 1. The choice of two valued elements is due to the simplicity and reliability of their hardware realisation. This does not represent any serious restriction in choice of number system because it is possible to encode other number systems into binary form.

A fundamental property of computer arithmetic is that it always deals with number representations of finite length. This means that the hardware result of a given operation may differ from the theoretical and that provision must be made for detection of overflow and underflow conditions.

We begin by discussing conversions from one form of number representation to another. Negative numbers and arithmetic operations are then considered with major emphasis on the binary system.

Positional number systems

Most commonly used number systems represent an integer as a string of digits each of which has position significance.

Thus in radix or base 10

$$789_{10} \equiv 7.10^2 + 8.10^1 + 9.10^{\circ}$$

In general terms for an m digit number N_r with digits d in radix r

$$N_r = d_o r^{m-1} + d_1 r^{m-2} \ldots\ldots\ldots\ldots d_{m-1} \qquad \ldots.(13.2a)$$

Thus

$$678_9 = 6.9^2 + 7.9^1 + 8.9^{\circ}$$

or

$$356_8 = 3.8^2 + 5.8^1 + 6.8^{\circ}$$

A polynomial form of representation is sometimes more convenient for evaluations.

$$367_{10} = ((3.10 + 6)\ 10 + 7)$$

$$456_8 = ((4.8 + 5)\ 8 + 6) \qquad \text{or}$$

in general form.

$$N_r = (--((d_o r + d_1)\ r + d_2)\ r ----+ d_{m-1}) \qquad \ldots.(13.2b)$$

Radix conversion

Conversion of integers from one radix representation to another is achieved by successive division. Thus to convert an integer from radix 10 to radix 8 we first divide by 8; the remainder is the least significant digit. The operation is then repeated.

Example:

Convert 367_{10} to base 8

$$8_{10}\overline{)\,367_{10}}$$

$$8_{10}\overline{)\,45} \qquad \text{and remainder } 7$$
$$\phantom{8_{10})\,}5 \qquad \text{and remainder } 5$$

Thus $367_{10} \equiv 557_8$

This corresponds to the reverse of the process of evaluation in nested form, the principle being apparent in (13.2b). Thus the remainder after the first division by r is d_{m-1} and after the second division d_{m-2} and so on.

Non-integer numbers are conventionally represented in two parts, an integer part and a fraction part. For example –

$$62.159 = 6.10^1 + 2.10^{\circ} + 1.10^{-1} + 5.10^{-2} + 9.10^{-3}$$

It is convenient to treat fractions separately and to represent a fraction of i places as –

$$F_r = f_o r^{-1} + f_1 r^{-2} + f_2 r^{-3} + --------- f_i r^{-(i+1)} \qquad \ldots.(13.2c)$$

or in nested form as –

$$F_r = r^{-1}\ (f_o + r^{-1}\ (f_2 ----- r^{-1}(f_i)----) \qquad \ldots.(13.2d)$$

The problem is to find f_o then f_1 and so on. This is achieved by multiplication by r which yields f_o as the integer part and a fraction part which is again multiplied by r to derive f_1 and so on. Continuation of the process will yield the elements of f as far as is desired.

Example:

Convert 0.367_{10} to 4 places in base 8

$$\begin{array}{r} 0.367 \\ \underline{\times\ 8} \\ \underline{2}.936 \\ \underline{\times\ 8} \\ \underline{7}.488 \\ \underline{\times\ 8} \\ \underline{3}.904 \\ \underline{\times\ 8} \\ \underline{7}.232 \end{array}$$

Thus to 4 places $0.367_{10} \equiv 0.2737_8$

It is important to note that exact conversion of fractions may not be possible.

Representation of negative numbers

The representation of negative numbers may take various forms the choice of which may be determined by considerations such as the simplicity of hardware realisation. For example, it may be desirable to avoid building a complete subtraction mechanism as well as an addition mechanism. Practical mechanisms are also finite in length. Thus, for example, a mechanism capable of working to two decimal digits can be represented as a circle (fig. 13.2a) from which it is clear that the *maximum* and minimum numbers are adjacent.

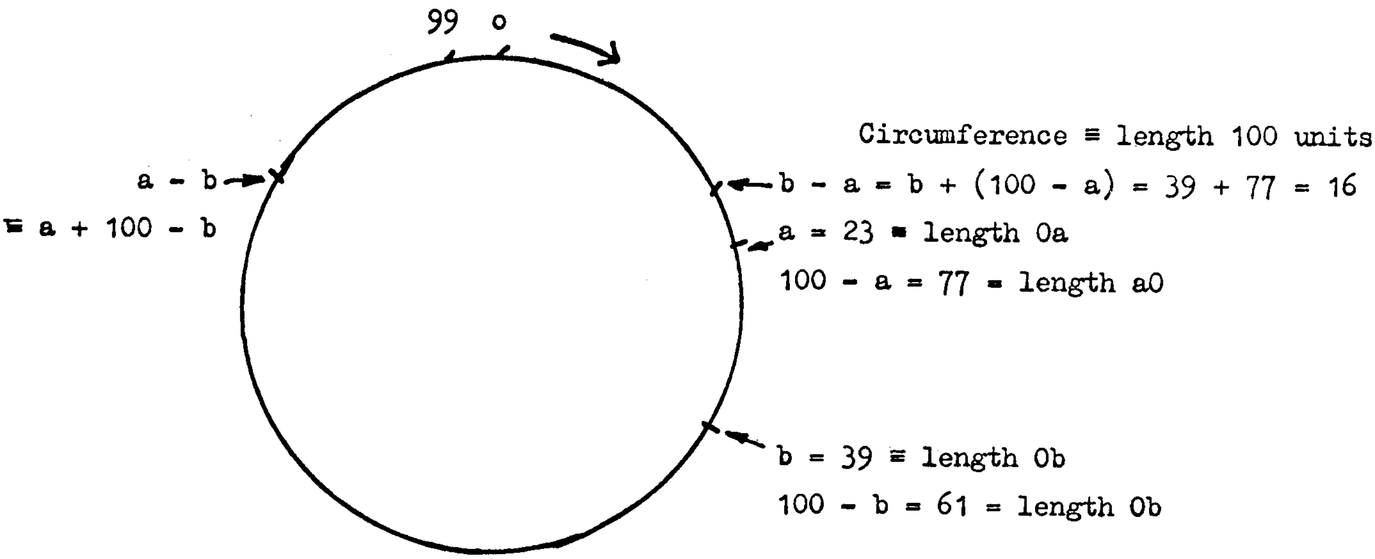


FIG. 13.2a Representation of Negative Numbers in Complement Form

A given number is then represented as a portion in the circumference, and addition corresponds to clockwise movement. Thus, if subtraction or anti-clockwise motion is not allowed it is apparent that we can always get to the correct result position by addition of the appropriate quantity. Thus, to form $a - b$ we can instead do the addition

$$a + (100 - b)$$

This is the basis of the complement system of representing negative numbers and leads to the general rule that addition of the complement is equivalent to subtraction. Formation of the complement is simply achieved by subtracting each digit from 9 and adding one into the least significant position. The validity of this lies in the observation that

$$100\text{---}0 = 99 \text{---} 9 + 0 \text{---} 01$$

and it is particularly simple to mechanise since no "borrow" digits can occur.

Examples:

If a = 39 b = 23

$$a - b \equiv a + 100 - b = 39 + 77 = 16$$

$$b - a = 23 + 61 = 84$$

The result 84 is the representation in complement form of −16. The sign digit may be automatically included by writing a positive number as od_1d_2 ---- and negative as $1d_1d_2$ --- and allowing the rules of binary addition to apply to the sign position. Thus in repeating the examples –

$$a - b = 039 + 177 = 016$$

$$b - a = 023 + 161 = 184$$

This form of complement is called the radix complement – in this case the tens complement. An alternative is radix minus one or nines complement. Thus the nines complement of 23 is 99 − 23 = 66. However, it is also apparent from the illustration that the number zero has two alternative forms, 99 and 00. It is also apparent that the correct result for b − a requires correction digit.

$$b - a = 39 - 23 = 16 = 39 + (99 - 23) + 1$$

The correction digit is automatically supplied if the carry from the most significant place is allowed to enter the least significant digit position. This is known as end-around carry and is necessary in all (radix −1) systems

$$\begin{array}{r} 039 \\ \underline{176} \\ (1)\,015 \\ \underline{(1)1} \\ 016 \end{array}$$

Although our discussion of negative representations has concentrated on the decimal system the same reasoning is applicable to numbers in other bases. We now give more detailed discussion of binary arithmetic since this is the basis of most design practice.

Binary arithmetic

The rules of binary arithmetic are greatly simplified by the fact that only two symbols are possible, 0 and 1 and were demonstrated in the section covering switching circuits and logic. The rules are summarised below:–

Addition

The input digits are the operands x and y and the input carry c. Table 13.1b is the truth table for binary addition.

Negative numbers

The three main alternatives correspond to those of the decimal system; sign and magnitude, two's complement and ones complement. In the following examples the sign digit is at the left and separated by a comma. We show the three alternative forms of negative representation:–

Positive representation of 13 is		0,1101
Negative representation in sign magnitude is		1,1101
" "	in ones complement	1,0010
" "	in two's complement	1,0011

The sign digit is treated identically with the number digits in addition or subtraction and end around carry must be allowed with ones complement. The following examples illustrate the principles involved and show the decimal equivalent:–

Number	Sign Magnitude	Ones	Two's
–13	1,01101	1,10010	1,10011
–27	1,11011	1,00100	1,00101
–15	1,01111	1,10000	1,10001

Two's complement

$$+27 - 15 \equiv \begin{array}{r} 0,11011 \\ +\ \underline{1,10001} \\ 0,01100 \end{array} \equiv +12$$

$$-13 + 27 \equiv \begin{array}{r} 1,10011 \\ +\ \underline{0,11011} \\ 0,01110 \end{array} \equiv +14$$

$$-27 + 15 \equiv \begin{array}{r} 1,00101 \\ +\ \underline{0,01111} \\ 1,10100 \end{array} \equiv -12$$

Ones complement

$$+27 - 15 \equiv \begin{array}{r} 0,11011 \\ +\ \underline{1,10000} \\ 0,01011 \\ \underline{1} \\ 0,01100 \end{array} \equiv +12$$

$$-13 + 27 \equiv \begin{array}{r} 1,10010 \\ +\ \underline{0,11011} \\ 0,01101 \\ \underline{1} \\ 0,01110 \end{array} \equiv +14$$

$$-27 + 15 \equiv \begin{array}{r} 1,00100 \\ +\ \underline{0,01111} \\ 1,10011 \end{array} \equiv -12$$

Multiplication

Basic rule

0 x 0 =0

0 x 1 = 1

0 x 1 = 1

Example:

$$13 \times 27 \equiv \begin{array}{r} 01101 \\ \times\ \underline{11011} \\ 01101 \\ 01101 \\ 00000 \\ 01101 \\ \underline{01101} \\ 101011111 \end{array} \equiv 351$$

Division

The steps are analogous to those in decimal arithmetic. A subtraction is allowed if the remainder is positive

Example:

450 ÷ 15 = 30

```
01111)111000010(11110
      01111
      011010
       01111
       010110
        01111
        001111
         01111
         000000
```

Octal is a convenient way of working with binary numbers. Numbers form groups of 3 bits which are given octal symbols 0 to 7. The chances of human error are reduced in this way.

Hexadecimal uses groups of 4 bits and the symbol set 0 to 9 A B C D E F . Thus a binary computer with a word length of 36 bits could represent bit patterns by 12 octal digits or nine hexadecimal digits. This is commonly done to give a more readable form than pure binary.

Exercise: Repeat the previous examples of binary arithmetic in octal and hexadecimal arithmetic.

13.3 Basic computer organisation

Registers

The circuits that store information in a computer can be divided into two classes: registers and memory cells. Register circuits are combined with logic circuits to build up the arithmetic, control and other information processing parts of the computer. The term "memory" is commonly reserved for those parts of a computer that make possible the general storage of information such as the instructions of a program, the information fed into the program, and the results of computation. Our procedure will be firstly to discuss arithmetic logic, then memory devices and systems and finally the organisation of a complete computer.

Registers are usually made up of one-bit storage circuits called flip-flops. A typical flip-flop circuit has two input terminals, set and reset, and two output terminals to provide both signal and complement. A flip-flop is usually named and may be represented symbolically as shown in fig. 13.1a. Thus, if a signal pulse is applied to the set terminal A becomes 1 and remains so until a pulse is applied to the reset terminal. Application of simultaneous pulses to both set and reset is not allowed and the circuit can be regarded as remembering its most recent input.

Arithmetic logic

Logic circuits and registers can be combined to perform elementary arithmetic operations. Table 13.1b is a truth table for one digit binary addition in which the inputs to the adder are the binary digits X and Y combined with the "input carry" The outputs are the sum digit and the "carry out". Fig. 13.1b is an implementation of this truth table based on AND, OR circuits.

In a computer employing binary arithmetic the basic number length could be say 32 bits. Numbers of such length can be added in two general ways. One way is to use an adder for each bit assembled as shown in fig. 13.3c. The inputs to this adder are the two four bit numbers: $X_3\ X_2\ X_1\ X_0$ and $Y_3\ Y_2\ Y_1\ Y_0$ and the output a five bit sum: $S_4\ S_3\ S_2\ S_1\ S_0$. The inputs to the adder can be provided by two four bit registers and the sum can be set into a five bit register following reset of the register to zero.

The second method of addition is serial which requires presentation of the appropriate bit pairs in correct sequence. One way of achieving this is to use special registers called shift registers which have the ability to shift information from one stage to the next on application of an input to the shift terminal. The essential difference is that only one binary adder is used and that the "carry out" digit is used to set the state of the flip-flop which then forms the carry in for the next pair.

Memory systems

The speed and cost of a modern computer is largely determined by the speed and capacity of the storage or memory unit. As a result a great deal of effort has been devoted to the development of memory systems and to the improvement of memory devices.

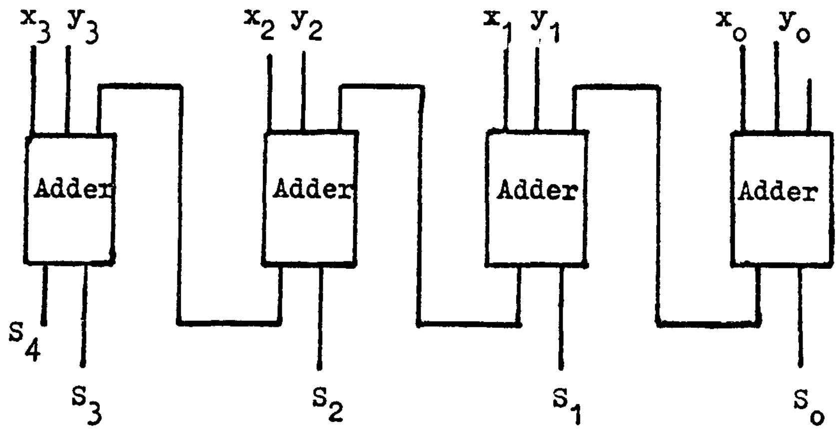


FIG. 13.3c Four Bit Parallel Binary Adder

Functionally a memory unit is a device which can read or write data into a location identified by an address. In the common situation each location holds a single word of data and the address is given by a number. An alternative form of address scheme is "content addressing" in which access is determined by the content of the word being sought. Although this form of addressing has many advantages it is more costly to mechanise and will not be further considered.

The important characteristics of a memory are access time, information transfer rate, capacity and cost. Table 13.3a is a comparison summary of different systems based on these characteristics. The range in each of the characteristics should be noted, in access time, capacity, and in cost per bit. Some of the most difficult problems in computer design are connected with proper use of storage – the store should appear to the user to have large capacity and rapid access but for economic reasons this must be realised by a hierarchy of storage devices.

Type of Memory	Access Time (microseconds)	Information Transfer Rate (Bits/second)	Capacity (Bits)	Cost (Dollars/bit)
Integrated circuit	10^{-2}	$10^9 - 10^{10}$	$10^3 - 10^4$	1
Typical Core or Film	0.5	10^8	10^6	10^{-1}
Large or Extended Core	5	10^8	10^7	10^{-2}
Magnetic drum/disc	10^4	10^7	$10^7 - 10^8$	10^{-3}
Tape Loop or Card	$10^5 - 10^6$	10^6	$10^9 - 10^{10}$	10^{-4}
Photographic	10^7	10^6	10^{12}	10^{-6}

Table 13.3a

Comparison of Memory Systems

The essential principles of the above systems are worth noting.

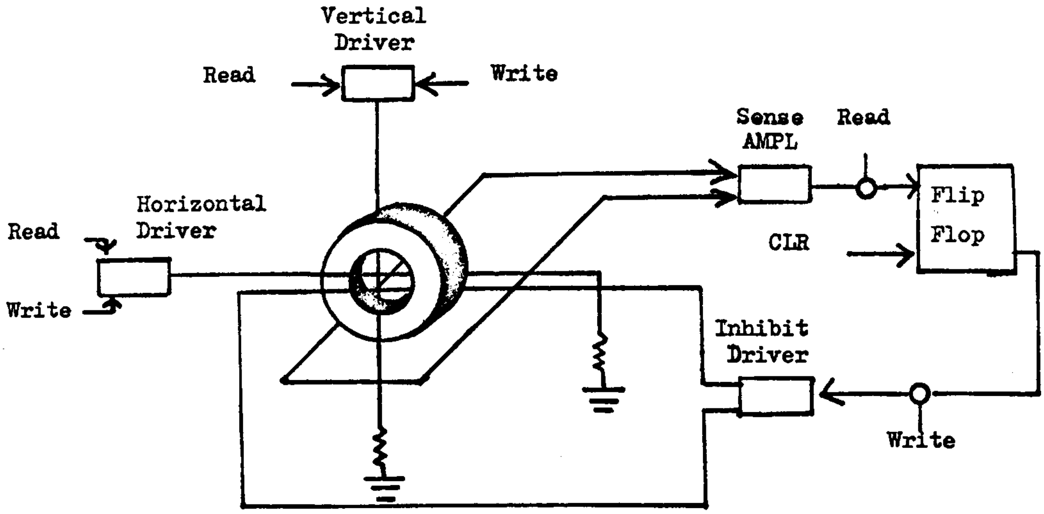


FIG. 13.3d A Single Memory Core with Associate Logic

The cores are made of ferrite with a rectangular hysterisis loop. Operation involves switching the direction of magnetisation between clockwise ('1' state) and anticlockwise ('0' state). Reading and writing signals each carry half the current needed to switch. During read, coincidence of the vertical and horizontal half currents causes switch of a core in the '1' state and an output in the sense amplifier, the '0' state produces no output. During write the horizintal and vertical currents are reversed which writes a '1' unless an opposing half current is in the inhibit wire in which case '0' state remains.

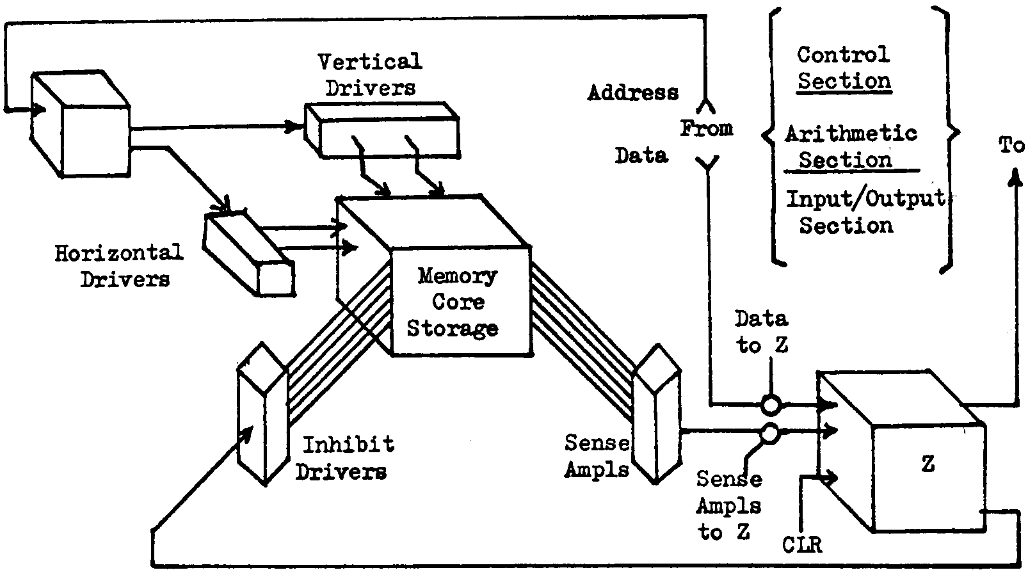


FIG. 13.3e A Complete Core Storage Unit

Such a unit may consist of 50 planes each containing 128 x 128 cores (16,384 words of 50 bits).

Advances in integrated circuits now permit hundreds of circuit elements to be packed into an area of the order of a tenth of an inch square. A typical high speed chip contains 16 flip-flops with an access time of less than 7.10^{-9} seconds (nanoseconds). A second technique (MOS.) is capable of much higher component density but results in slower circuits ($0.2.10^{-6}$ seconds). There is little doubt that these components will have an important role in future computers.

Magnetic-core memories are the most widely used random access memory systems. A normal system consists of a three dimensional array of about a million tiny magnetic cores or rings each of which can store one bit of information. The cores are made of ferrite with a rectangular hysteresis loop so that a 1 or a 0 can be stored as a clockwise or anticlockwise magnetic state. The rectangular characteristics also means that the direction of magnetisation cannot be changed by a current below the threshold (say 200 ma) but can be safely changed by double this current (400 ma). This threshold characteristic enables lower cost coincident current techniques to be used for addressing and writing into a memory stack. Fig. 13.3e and Fig. 13.3d illustrate in further detail the principles involved.

The main attraction of magnetic films as an alternative to cores is in their greater speed potential. However, there are problems in fabrication and in driving circuits.

Recording information on a moving magnetic surface is one of the oldest and more obvious methods of storing information. In recent years quite significant progress has been made in improving the design and quality of recording heads and surfaces. Various mechanical arrangements are used – drums, fixed disks, removable disks, magnetic tape, magnetic cards etc. all of which are designed to provide a range of performance capacity and cost parameters.

Photographic techniques have advantages where read-only access is required to high volumes. Some interesting proposals have also been made in the use of lasers.

Arithmetic unit

The arithmetic unit is responsible for execution of the logical and arithmetic operations. We will consider the organisation of a simple unit capable of fixed point addition, subtraction, multiplication and division. This demonstrates most of the important principles involved and is of historical interest since it is similar in organisation to the early machines from which modern machines have evolved. Fig. 13.3f shows a structure for an arithmetic unit using 4 bit words (a practical system may use 32 bits).

Data is transferred from the upper to the lower registers directly or via the adder unit and returned from the lower to upper with a right shift of one place. The example illustrates the individual steps involved in forming the product 1001 x 1101. The bits of the multiplier are sensed successively commencing with the least significant bit and used to determine whether the multiplicand should be added (m = 1) to form a new partial product. Shifting takes place on return from the lower to upper registers. It will be seen that the growing partial product successively displaces the individual bits of the multiplier thus achieving economy in register accommodation. The essential differences between this machine method and the normal hand method is that partial products are formed for each step rather than leaving addition until the end and that the position at which addition takes place is fixed so that the partial product is moved instead to give the same relative shift.

If we allow either the register value or the bitwise complement to enter the adder then subtraction is also possible. For example, if two's complement is used then A – D may be done by using the bitwise complement of D and injecting a one into the carry input at the least significant bit position.

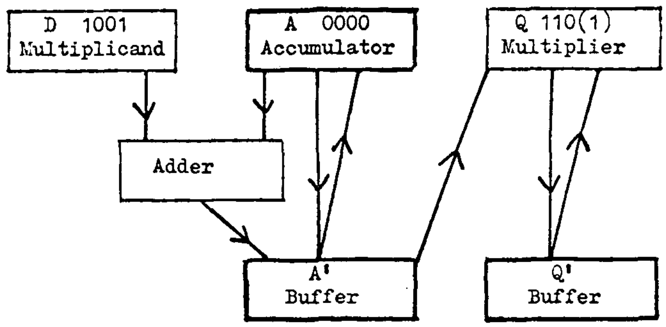
Division is possible if we include a path for shift left. In this case the process is one of trial subtraction at the end of which a left shift is done and the corresponding quotient digit placed in the least significant bit portion of Q. The process is illustrated in Fig. 13.3g.

The arithmetic structure so far developed will allow addition, subtraction, multiplication, division, and with minor additions shift right, shift left and the logical operations AND, OR, and EQUIVALENCE. It is therefore sufficient to execute an adequate set of operations for a general purpose fixed point binary computer. However, generation of the sequence of transfer operations which comprise an arithmetic operation such as multiplication requires a further unit called a control unit. In the next section we outline the organisation of a simple computer based on these concepts.

Organisation of a simple computer.

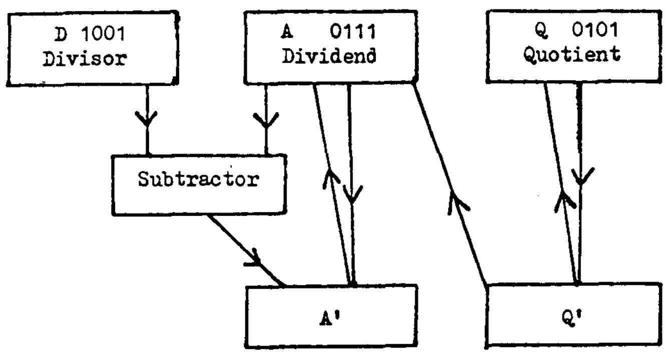
Fig. 13.3h shows the organisation of a simple general purpose computer.

The storage unit holds both instructions and data. it is accessed by supplying the address to the address register and then passing data to or from the store via a buffer register known as the data register. Data and instructions are initially loaded into the store by the input unit and results taken from the store by the output unit. This allows the arithmetic unit to derive program instructions and data from the store and thus to avoid being bound to peripheral unit speed. The basic operation cycle is in two phases as follows:–



AQ	0000	110(1)	Sense m = 1
A'Q'	1001	110(1)	Add multiplicand
AQ	0100	111(0)	Shift right 1. Sense m = 0
A'Q'	0100	111(0)	Copy
AQ	0010	011(1)	Shift right. Sense m = 1
A'Q'	1011	011(1)	Add multiplicand
AQ	0101	101(1)	Shift right. Sense m = 1
A'Q'	1110	101(1)	Add multiplicand
AQ	0111	0101	Shift right

FIG. 13.3f. Binary Multiplication in a Simple Arithmetic Unit



AQ	0111	0101	
A'Q'	0111	0101	Transfer
AQ	1110	1010	Shift left
A'Q'	0101	1010	Subtract divisor
AQ	1011	010(1)	Shift left, insert quotient bit = 1
A'Q'	0010	010(1)	Subtract divisor
AQ	0100	10(1)(1)	Shift left, insert quotient bit = 1
A'Q'	0100	10(1)(1)	Transfer, (subtraction goes neg.)
AQ	1001	0(1)(1)(0)	Shift left, insert q = 0
A'Q'	0000	0(1)(1)(0)	Subtract
AQ	0000	(1)(1)(0)(1)	Shift left. Quotient in Q

FIG. 13.3g. Binary Division in a Simple Arithmetic Unit

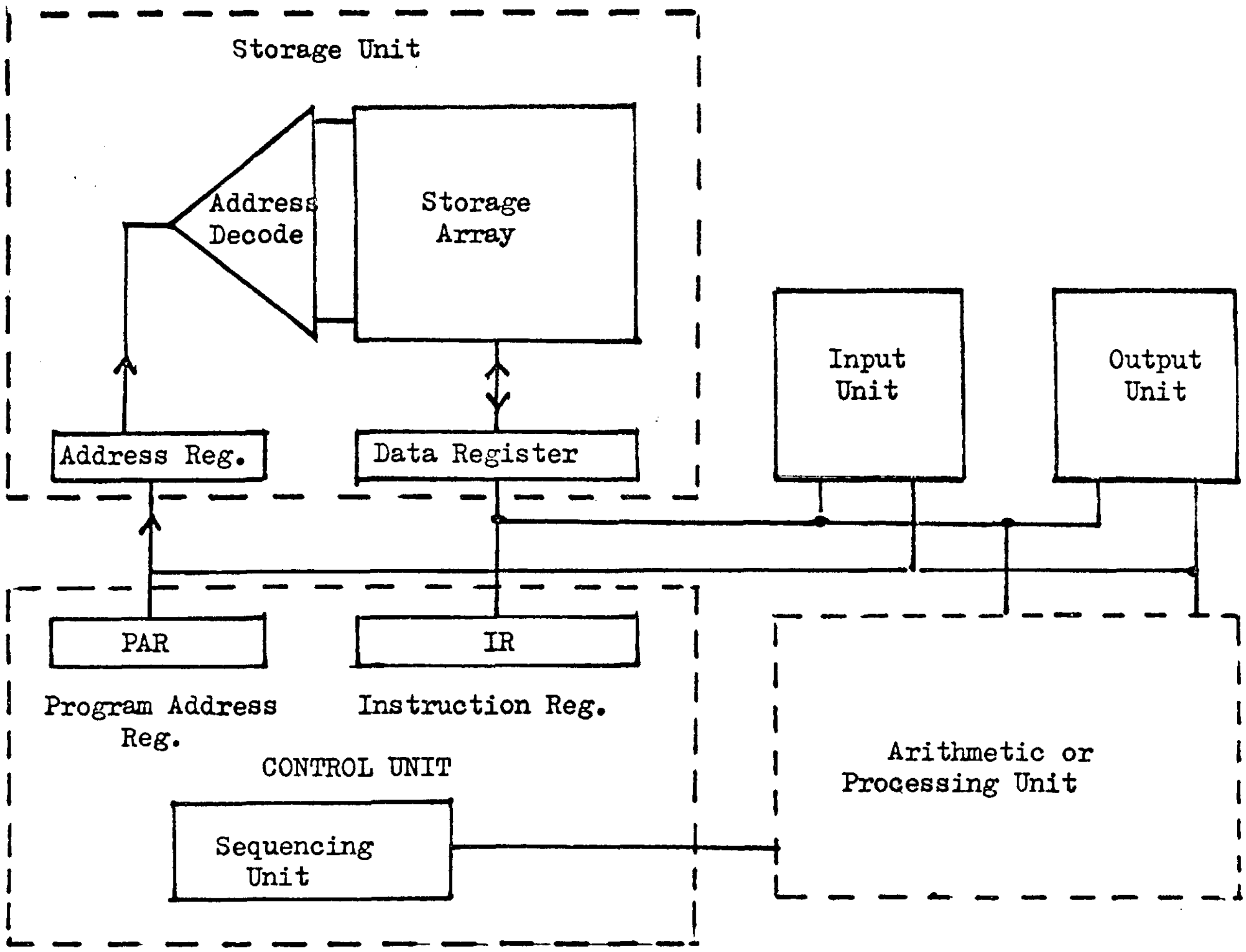


FIG. 13.3h Organisation of a Simple Computer

(1) *Read next instruction.* The location of the next instruction is the address held in the program address register. This address is transferred to the store address register and the contents of the location passed via the store data register to the instruction register where it sets up decoding circuits.

(2) *Execute instruction.* The instruction held in the instruction register is decoded and initiates the appropriate control sequence for execution. In general this will require another access to the store and in this case the store address is derived from the instruction register and the data word is passed to or from the arithmetic unit. Before returning to the "read next instruction" phase the address in the program address register is updated – in the usual case this consists of adding unity to the last value.

In the next chapter these concepts will be illustrated by specifying a machine and instruction set to be used in programming examples.

13.4 Exercises

1 Sketch a logic circuit to realise the carry output in a binary adder. Can you suggest more than one solution.

2 Parity checks are often used for error checking. The scheme involves addition of an extra bit to cause the number of ones to be even (even parity) or odd (odd parity). For example odd parity on the following 3 bit numbers would appear as –

Digits	Parity
000	1
001	0
010	0
011	1
100	0
101	1
110	1
111	0

Devise a circuit to check whether parity is correct in a given group. Comment on the effectiveness of this method of error detection.

3 Convert the following:–

157_{10} to binary

133_9 to octal

567_8 to base 9

110110_2 to decimal

4 Carry out the operations of addition, subtraction and multiplication on the following number pairs –

1011_2,	1001
56_8,	35_8
23_9,	38_9

5 Using the number pairs in 13.4 repeat the operation of subtraction using the radix complement and radix –1 complement.

13.5 Suggested Reading List

(1) "Switching Circuits for Engineers". M.P. Marcus (Prentice Hall).

(2) "Digital Computer System Principles". H. Hellerman (McGraw Hill).

(3) "Scientific American" Sept. 1966. This is a special computer issue.

Chapter 14

Introduction to Programming

NO, THE FUTURE offers very little hope for those who expect that our new mechanical slaves will offer us a world in which we may rest from thinking. Help us they may, but at the cost of supreme demands upon our honesty and our intelligence. The world of the future will be an even more demanding struggle against the limitations of our intelligence, not a comfortable hammock in which we can lie down to be waited upon by our robot slaves.

NORBERT WIENER
God & Golem, Inc., 1964.

Introduction to Programming

Programming is the precise specification of a sequence of operations designed to obtain a solution to a given problem. Programs may be written in various languages which are designed to suit different application areas (e.g. FORTRAN, ALGOL, COBOL, PL/1). There is no question that applications programs are best written in such languages. However, the object of this chapter is to acquaint the reader with the fundamentals of programming at a level where there is close correspondence between each instruction written and the corresponding hardware action of the computer. Thus we adopt the historical approach commencing with absolute address coding through symbolic assembly coding and leading in Chapter 15 to some basic concepts in the design of language processors such as FORTRAN. The emphasis is therefore on what is happening "under the bonnet" rather than in achieving proficiency in applications programming in a particular high level language. This is a skill which is largely acquired in other courses and is better appreciated following some experience in machine language programming.

The machine code chosen for use in examples is that of a hypothetical computer. However, a simulator program enables programs to be run as if a real machine was available. We will refer to the hypothetical computer as SAC being an abbreviation for Simulated Automatic Computer.

14.1 General description of SAC

SAC may be classified as a one-address, fixed word length, stored program decimal digital computer. The aim in this decision is to avoid complexity but retain the necessary programming features. A reasonable alternative choice would be a binary machine but this would either require the user to program in binary or assume the availability of conversion routines. We begin by describing the basic structure of the machine and then introduce instructions and new facilities as the need for them is demonstrated. (Fig. 14.1a).

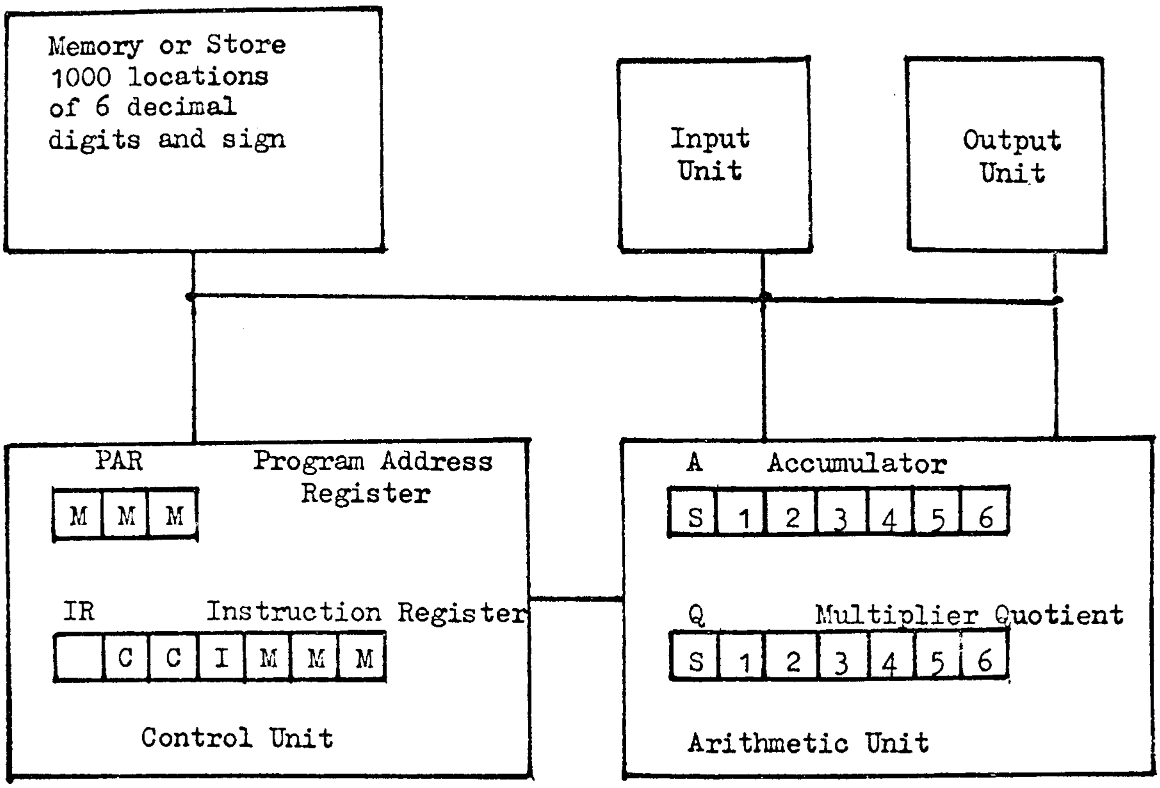


FIG. 14.1a Basic Structure SAC

Memory Unit

The memory has a capacity of 1000 words each of which is specified by a three digit location address in the range 0 to 999. Each word contains a sign and six decimal digit positions.

Arithmetic unit or central processing unit

The arithmetic unit uses two one word registers, an accumulator (A), and a multiplier – quotient register (Q). For certain instructions A and Q are linked and may be regarded as holding a 12 digit number with sign.

Control unit

The control unit uses two registers – an instruction register (IR) which holds the instruction to be executed, and the program address register (PAR) which holds the address from which the next instruction is to be read.

Input unit

This allows data to be read from an input device into a specified memory location. We will assume punched cards as the input medium and details of the instructions are given later.

Output unit

This allows data to be transferred from a given memory and printed on the output unit.

Instruction format

Each instruction occupies one word and is coded as six decimal digits. The first two digits with preceding sign (1 and 2 denoted CC) specify the function or operation to be performed e.g. addition, multiplication. The third digit (I) has a special function to be dealt with later. The remaining digits (4, 5, 6 denoted MMM) have several meanings depending on the function or operation part of the instruction. Typical interpretations are:–

(i) A location in memory.

(ii) The location address of the next instruction to be executed.

(iii) The number of digit positions that the contents of A or Q are to be shifted.

1	2	3	4	5	6
C	C	I	M	M	M

Fig. 14.1b Instruction Format.

We will now consider the instructions available with SAC using a classification based on function. At the same time three letter mnemonics will be defined which will be used in preference to numeric codes in all examples. We therefore assume availability of a program to convert to numeric codes. We will use the following initial values in the illustrative examples for each instruction and adopt a convention of brackets for "contents of storage location"

A	=	+ 001234
Q	=	+ 223344
(100)	=	– 007899
(101)	=	+ 000122
(102)	=	+ 000006

Information transfer instructions

Information may be transferred from register to register, memory to register or register to memory. The action is to copy the information source into the destination. This leaves the source unchanged but overwrites the previous content of the destination.

TMA

(CC = +10) Transfer the contents of memory location MMM to accumulator A.

Example:	TMA 101			
Before	A = +001234	(101)	=	+000122
After	A = +000122	(101)	=	+000122

TMQ

(CC = +11) Transfer the contents of memory location MMM to Q.

Example:	TMQ 101			
Before	Q = +223344	(101)	=	+000122
After	Q = +000122	(101)	=	+000122

TAM

(CC = +12) Transfer the contents of A to memory location MMM.

Example:	TAM 100			
Before	A = +001234	(100)	=	–007899
After	A = +001234	(100)	=	+001234

TQM

(CC = +13) Transfer the contents of Q to memory location MMM.

Example:	TQM 100			
Before	Q = +223344	(100)	=	–007899
After	Q = +223344	(100)	=	+223344

TAQ

(CC = +44) Transfer contents of A to Q.

Example:	TAQ			
Before	A = +001234	Q	=	+223344
After	A = +001234	Q	=	+001234

TZA

(CC = +01) Transfer zero to A.

TQA Transfer contents of Q to A.

Example:	TQA			
Before	A = +001234	Q	=	+223344
After	A = +223344	Q	=	+223344

In the preceding examples all transfers involved a full word. The following instructions involve only the address part of the word.

SAA

(CC = +14) The address part of the word in memory location MMM is stored in A.

Example:	SAA 100			
Before	A = +001234	(100)	=	–007899
After	A = +001899	(100)	=	–007899

SAM
(CC = +15) The address part of the word in A is stored in memory location MMM.

Example:	SAM 100			
Before	A = +001234	(100)	=	–007899
After	A = +001234	(100)	=	–007234

Arithmetic instructions

Two classes of arithmetic instructions are allowed – integer and floating. With integer arithmetic the decimal point is understood to lie at the right of position 6. If this was the only form of arithmetic allowed then in many cases we would have to scale quantities. This is a programming burden so that we could introduce floating arithmetic which greatly increases the allowed range of numbers. However, we do not require floating arithmetic in the examples chosen for this chapter.

ADD
(CC = +02) The contents of memory location MMM are algebraically added to the contents of A leaving the sum in A.

Example:	ADD 101			
Before	A = +001234	(101)	=	+000122
After	A = +001356	(101)	=	+000122

SUB
(CC = +03) Subtract the contents of memory location MMM from the contents of A leaving the difference in A.

Example:	SUB 101			
Before	A = +001234	(101)	=	+000122
After	A = +001112	(101)	=	+000122

MUL
(CC = +04) Multiply the contents of Q by the contents of memory location MMM leaving the 12 digit product in AQ.

Example: MUL 102

Note that the initial value in A is destroyed.

Example:	MUL 101			
Before	A = +001234	Q	=	+223344
		(101)	=	+000122
After	AQ holds +223344 x +000122 = +000027247968			

DIV
(CC = +05) Divide the combined contents of AQ taken as a 12 digit dividend by the content of memory location MM. The quotient to six digits is left in Q and the remainder is A. We will assume new values in A and Q.

Example:	DIV 102			
Before	A = +000000	Q	=	+000045
		(102)	=	+000006

After	A = +000003	Q	=	+000007
		(101)	=	+000006

Shift instructions

The contents of A or Q or AQ combined may be shifted right or left any number of decimal places. The amount of shift is coded as MM and in all cases the sign digit is not involved and vacated places are filled with zeros.

SAR
(CC = +20) Shift A right

SAL
(CC = +21) Shift A left

AQR
(CC = 22) Shift AQ right

AQL
(CC = +23) Shift AQ left

SQR
(CC = +24) Shift Q right

SQL
(CC = +25) Shift Q left

Jump instructions

The program address register (PAR) is normally incremented for each instruction executed so that instructions are obeyed consecutively. The jump instructions enable this sequence to be broken by setting PAR to a new value. An unconditional jump is provided and a number of conditional jumps.

JMP
(CC = +30) Jump to MM unconditionally. The value in the PAR is made MM and the next instruction taken from MMM.

Example:	JMP 305
Before	PAR = 201
After	PAR = 305

JAZ
(CC = +31) Jump to MMM if the accumulator is zero.

JAP
(CC = +32) Jump to MMM if A is positive.

JAQ
(CC = +33) Jump if the value in A equals that in Q.

JAN
(CC = +34) Jump if A is negative.

HLT – Halt. The computer is stopped.

14.2 **Programming examples**

A considerable amount of planning is needed before a computer is approached to solve a problem. A computer can only attempt to solve a problem when given a complete, correct, and unambiguous statement of the procedure to be followed.

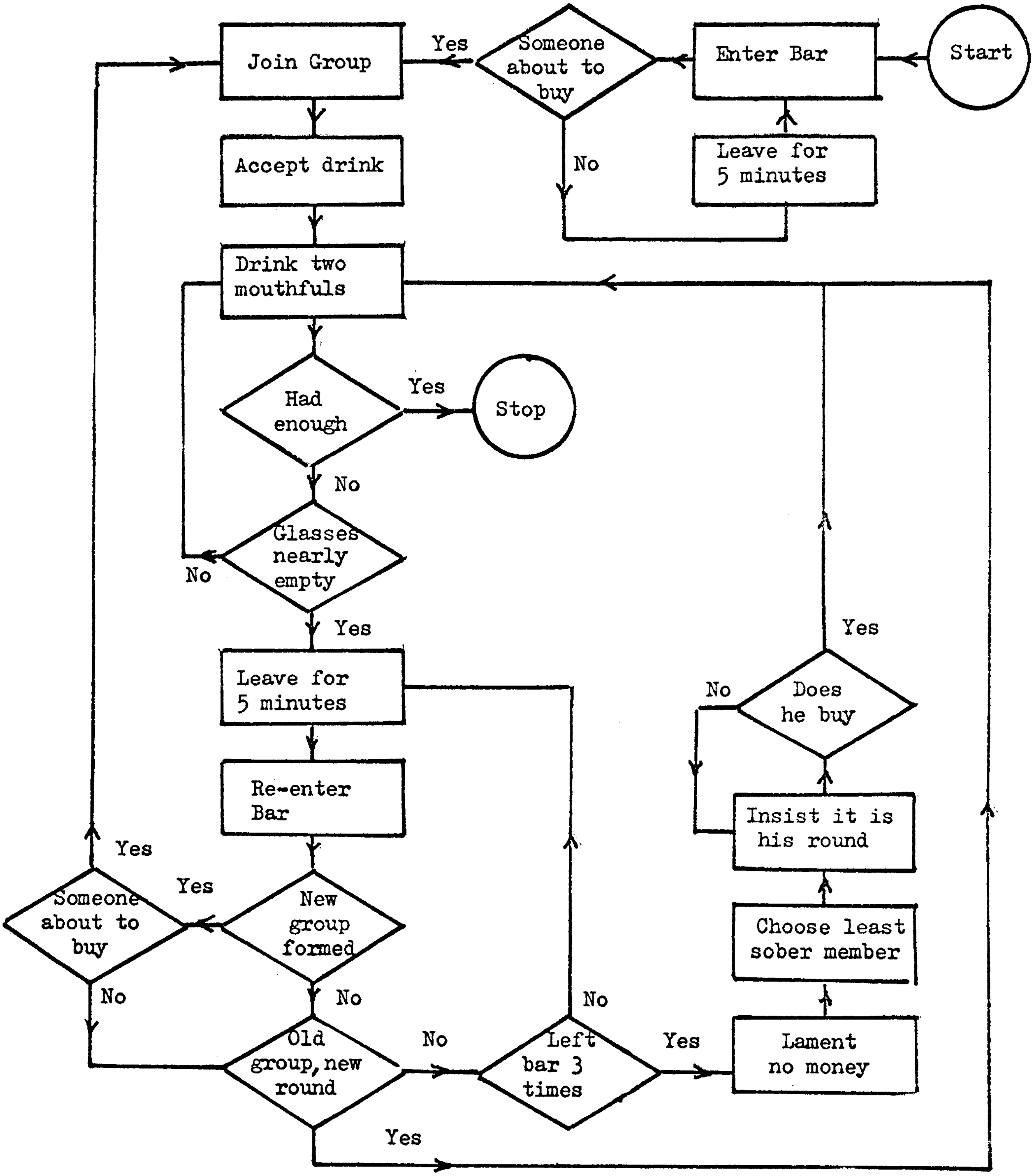


FIG. 14.2a Flow Diagram, Bar Strategy

Such a computing procedure is called an *algorithm.* The term algorithm derives from an ancient Arabic text "Kitub aljabra w'al-muqabula" by Al-Khowarizmi (c 825). This derivation was lost until quite recently and interpretation was varied. Some of the interesting interpretations were "any peculiar arithmetic method" and "painful arithmetic". The algorithms which interest us in computing are those which are capable of implementation as computer programs and have therefore a number of special features as follows:—

finiteness	:	an algorithm must always terminate after a finite number of steps.
definiteness	:	each step of the algorithm must be precisely defined, rigorously and unambiguously for all possible cases.
effective	:	an algorithm should be effective and efficient and state the conditions under which this is obtained.
input and output	:	in general an algorithm will require input of specified type and will produce an output which is the desired transformation of the input.

Flow charts are a graphical way of representing algorithms and are a very useful and universal way of expressing such procedures. We can draw flow charts for many procedures used in our daily lives and it is a useful and helpful method. Fig. 14.2a is intended to illustrate this point. We adopt the following conventions:—

(1) Circles are used for locating start, stop, and links to other charts.

(2) Plain rectangles are used for statements describing some action to be followed.

(3) Angular ends are used when branching decisions are involved.

Direction of flow is generally from top to bottom. When a suitable algorithm has been chosen and expressed as a flow chart the problem is then ready for programming. This involves translation of the logic embodied in the flow chart into a specific set of instructions that will cause the computer to implement the procedure. The exact form that these instructions assume is dependent on the computer being used and the languages available. A modern computer installation normally supports many languages each of which is best suited to the expression of algorithms from different areas of application: examples are FORTRAN, ALGOL (evaluation of mathematical procedures), COBOL (commercial applications), SNOBOL (string and text data), LISP (linked lists), GPSS (simulation and modelling). We propose to deal in this chapter with the expression of algorithms in the language of the machine itself using the simple machine already described. This choice is made to reveal what is happening in the machine in the belief that this is the right approach to using higher-level languages. There is no doubt that application programs are better written in higher level languages.

Example: 14.1

Given three numbers X, Y, and Z in locations 300, 301, 302, find the largest of them and place it in location 500. Start the program at location 200.

14.2b Flow diagram

PAR	Instruction		
200	TMA	300	X
201	SUB	301	X – Y
202	JAN	206	Jump if X < Y
203	TMA	301	Y
204	SUB	302	Y – Z
205	JAN	212	Jump if Y < Z
206	TMA	300	X
207	SUB	302	X – Z
208	JAN	215	Jump if X < Z

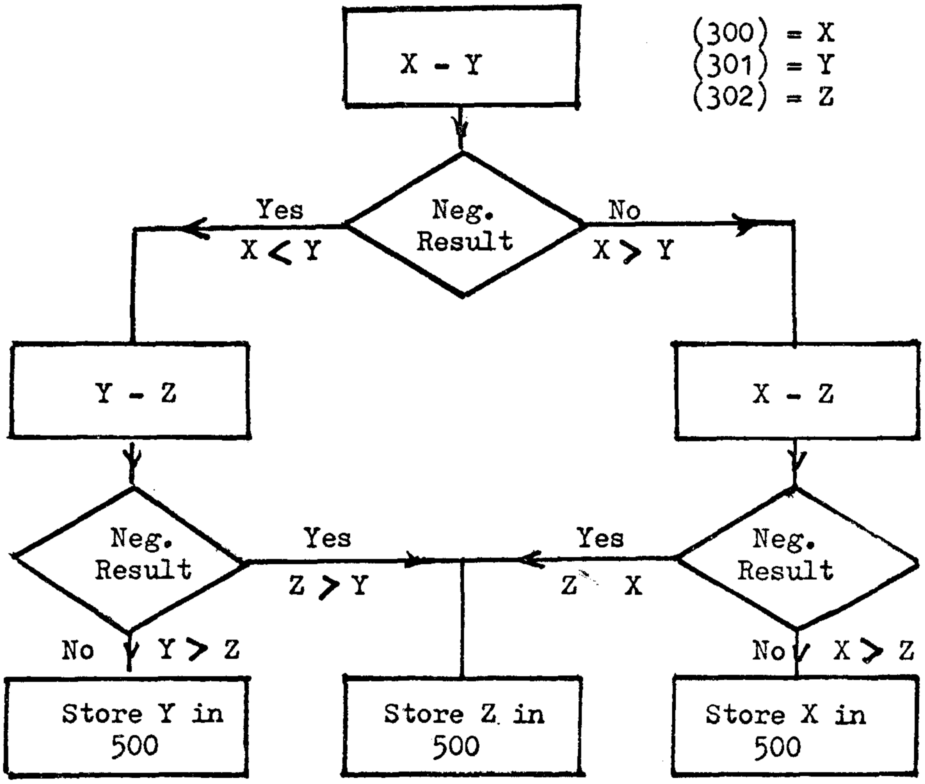


FIG. 14.2b Flow Diagram Example

209	TMA	300	X
210	TAM	500	Store X in 500
211	HLT		Halt
212	TMA	301	Y
213	TAM	500	Store Y in 500
214	HLT		Halt
215	TMA	302	Z
216	TAM	500	Store Z in 500
217	HLT		Halt

Program loops.

There are many situations in programming where we wish to repeat a group of instructions a number of times. We refer to these situations as program loops and they may arise for various reasons. For example, we may wish a given calculation to be repeated for several hundred different sets of data or we may wish to repeat the process until a given error condition is reached. Looping thus reduces the number of instructions we need to write, a program loop therefore forms a very important concept. Fig. 14.2c shows a flow chart for a generalised loop. In the examples which follow brackets are used to enclose address fields altered by the program.

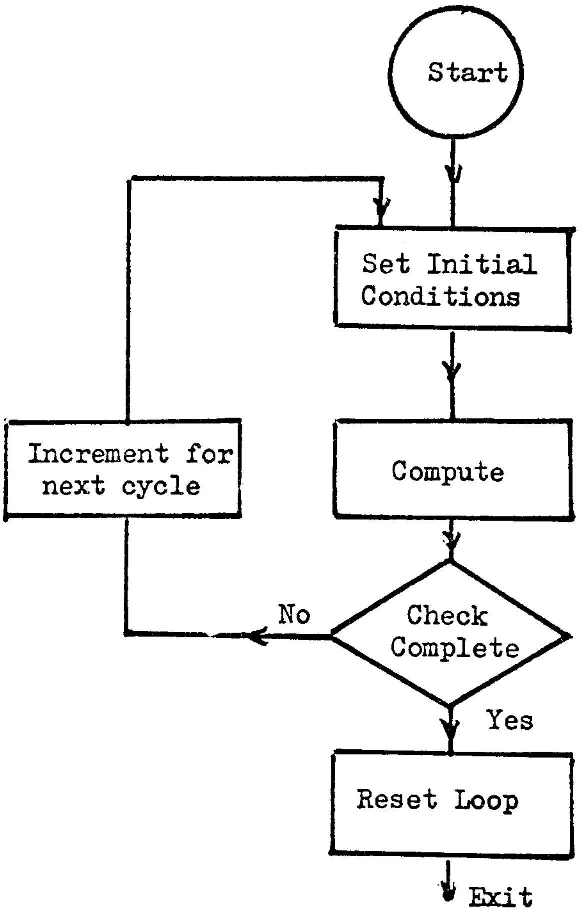


FIG. 14.2c   Flow Chart of Generalised Loop

Example: 14.2

Given 300 numbers stored in 200 through 499 inclusive form their sum and store the result in 650. Start the program at 100. Assume +1 is held in location 150 and 500 in 151.

PAR	Instruction		Comments
100	TZA		Initialise
101	TAM	650	
102	TMA	650	Compute. Values
103	ADD	(200)	are for first
104	TAM	650	cycle.
105	SAA	103	Increment
106	ADD	150	address
107	SAM	103	

108	SUB	151	Test for
109	JAZ	111	completion
110	JMP	102	
111	HLT		End

Example: 14.3

Locations 0 to 299 contain random single digit integers. Write a program to make a frequency count of how many times each digit occurs. (Store the results in consecutive locations commencing at 400 and start the program at 100. Assume the constant +1 in location 412, 400 in 413, 300 in 414 and locations 400 - 409 are initially cleared: –

PAR	Instruction		Comment
100	TMA	(0)	digit
101	ADD	413	digit + 400
102	SAM	104	address is
103	SAM	106	digit + 400
104	TMA	(400)	frequency count
105	ADD	413	+ 1
106	TAM	(400)	count + 1
107	SAA	100	calculate address
108	ADD	412	of next digit
109	SAM	100	
110	SUB	414	Test completion
111	JAZ	113	zero if complete
112	JMP	100	Next loop
113	HLT		Finish.

Example: 14.4

Compute the cubes of the odd numbered integers from 1 to 23 and place them in successive locations commencing at 200. Assume the digit 1 is held in location 99 and 2 in location 101, 23 in location 103 and start the program at location 1 300.

PAR	Instruction		Comments
300	TMA	99	Initial value
301	TAM	100	X = 1
302	TM~~A~~Q	100	Compute X
303	MUL	100	X^2
304	MUL	100	X^3
305	T~~A~~M G	(200)	200 is initial value

306	TMA	100	X
307	SUB	103	X - 23
308	JAZ	316	Test for completion
309	TMA	100	X
310	ADD	101	X + 2, Next value of X
311	TAM	100	(100) = X + 2
312	SAA	305	Increment the value
313	ADD	099	of address for
314	SAM	305	storing
315	JMP	302	
316	HLT		

Input-output instructions

We will assume the input medium is punched cards and the output medium also punched cards or alternatively a printer. The way in which data is represented on the card is different from that in the machine so that we will assume the input units can do the necessary conversions. We distinguish two types of data on cards –

(1) Decimal codes

(2) Alphanumeric codes i.e. letters and symbols plus figures.

The decimal digits have to be converted from the form used on the card to the correct internal 4 bit codes. Alphanumeric codes are converted into decimal digit pairs so that each word can hold 3 alpha codes. Although a card has 80 column our units ignore the last 5 columns which may be used for other purposes such as sequence numbers.

RDA

(CC = +52) Read card alphanumerically. Alphanumeric data punched in columns 1 – 75 are converted into two digit representations and stored 3 per word commencing in memory location MMM through MMM +24.

PTA

(CC = +55) Print alphanumerically. The 75 characters contained in memory locations MMM through MMM +24 inclusive are printed.

RDN

(CC = +50) Read card numerically. Numeric data or instructions are read in decimal form. Input cards are punched ten words per card starting in column 1 with a space between each signed six-digit quantity. Words are stored at the ten consecutive locations MMM to MMM + 9. A blank word will be stored as negative zero.

PTN

(CC = +51) Print numerically. The ten words in locations MMM to MMM + 9 are typed as a single 80 column line in a format identical to that of a numerically punched card.

The repeat instruction

We have demonstrated the value of program loops and experienced some of the "red tape" involved in constructing a well behaved loop. Care must be taken to initialise the loop to increment addresses correctly, and to terminate the loop at the right point. We now introduce a new machine instruction which greatly simplifies these steps.

RPT I

(CC = +99) Perform the next I instructions ($0 < I < 9$) a total of MMM cycles. The effective address for each instruction under repeat influence will be the original address the first time the instruction is executed, and there-after

will be I more than the preceding time for I in the range 0 – 5 and 10 – I less for I in the range 6 to 9. This is summarised below: –

Nominal I	Effective increment
0	0
1	1
2	2
3	3
4	4
5	5
6	-4
7	-3
8	-2
9	-1

In future examples the I field will be included in all instructions.

Example: 14.5

We illustrate the use of this instruction by repeating example 14.3. Note the reduction in the number of instructions required: –

PAR	Instruction			Comments
100	RPT	7	300	Repeat the next 7 instructions 300 times
101	TMA	1	000	Bring digit
102	ADD	0	413	400 + digit
103	SAM	0	105	Store computed
104	SAM	0	107	address
105	TMA	0	(400)	Add 1 to appropriate
106	ADD	0	413	digit counter
107	SAM	0	(400)	

The preceding instructions deal with data in blocks equal to a punch card or a line of printing. Sometimes it is desirable to deal with single characters and digits and have these input and output via a special operator console typewriter. We therefore introduce the following instructions: –

TID

(CC = +56) Type in digit. A and Q are shifted left one place and the digit entered as the right most digit.

TOD

(CC = +57) The left most digit of A is typed following which A and Q are shifted left one place.

TIC

(CC = +58) Type in character. A and Q are shifted left 2 digit positions and the character placed in the rightmost two digit positions of Q.

TOC

(CC = +59) Type out character. The character typed is the leftmost two digits of A. A and Q are then shifted left 2 positions.

Used in a repeat loop these instructions allow printing of up to 12 unsigned decimal digits or 6 characters.

Example: 14.6

Read a numeric card, form the sum of the ten numbers stored on it and print the result in the typewriter. Use locations 20 – 29 for input output and start the program at 300.

PAR	Instruction			
300	RDN	0	020	Card to 20 – 29
301	TZA	0	000	A = 0
302	RPT	1	020	Add the ten
303	ADD	1	020	numbers
304	RPT	1	007	Type out A
305	TOD			contents
306	HLT			

Sub-routines

A sub-routine is a code sequence for a useful sub-process. For example, there is a frequent need for functions such as square root, sine, cosine, log etc. Programs may be more easily written by making use of available sub-routines, and this may be done in two ways. We could simply imbed a copy of the sub-routine at each point in the main program where it is needed or we could provide just one copy of the sub-routine and provide some way for entering it from the main program and returning to the main program at the end of the sub-routine. The first method is called an open sub-routine and the second a closed sub-routine. The difference in concept is shown in Fig. 14.2d.

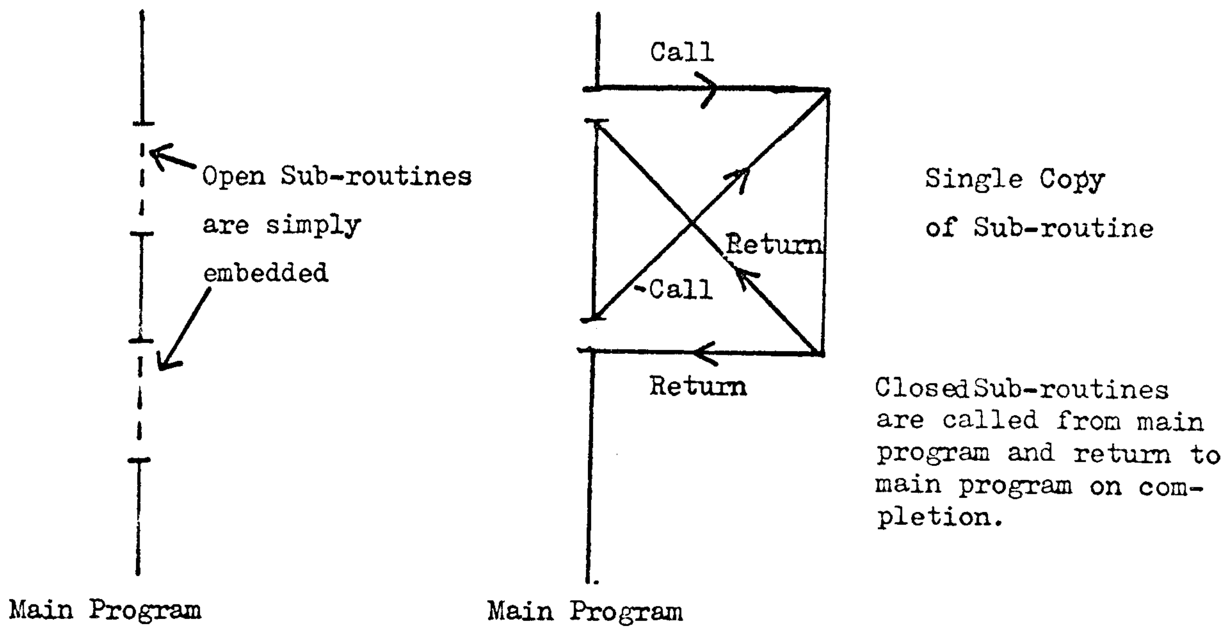


FIG. 14.2d Open and Closed Sub-routines

The closed sub-routine is more economic in storage but slightly longer and slower in execution because of the need to pass control and parameters between the main program and sub-routine. Open sub-routines are usually very short and their incorporation into a main program can be done automatically by the assembler – *macro instructions* are used to name the routines. Closed sub-routines are more difficult to use and it will be necessary to discuss some of the techniques for link and return, and for passing parameters. We are faced with three problems: –

(a) How to pass data from the main program to the sub-routine.

(b) How to return data from the sub-routine to the main program.

(c) How to return control to the correct point in the main program.

We introduce an additional hardware feature to assist in this problem. The feature consists of a register (Jump address register JAR). At the conclusion of a successful jump JAR holds the address of the next instruction in line after the jump instruction, i.e. the instruction which would have been selected next if the jump had not occurred. We provide an instruction to transfer the contents of JAR to A or to memory

TJA

(CC = +26) Transfer the contents JAR to A.

TJM

(CC = +27) Transfer the contents of JAR to the address portion of memory location MMM.

There are various conventions which could be developed using the above instructions.

Example: 14.7

Show linking and return from a main program at 200 to a sub-routine at 300.

```
200 JMP 300
201                299 JMP *          * becomes 201 after
                   300 TJM 299          instruction 300

                       JMP 299
                       Sub-routine

Main Program
```

Note that this is only one of a number of possible conventions that could be used.

Assembler programs

The type of coding we have followed assigns both data and instructions to pre-selected memory locations. The only concession we have made is to use mnemonic symbols instead of function codes. There are two apparent difficulties; firstly programs are difficult to modify since a single change may affect a whole group pf orders or data; secondly if we have two programs written independently it could be quite difficult to combine them if their use of memory overlaps.

The answer to these problems is to avoid the use of absolute addresses and to use symbolic addresses to name locations in the program. We are therefore assuming the existence of a program which will –

(a) Translate mnemonic function codes to actual codes.

(b) Replace symbolic addresses with absolute addresses.

Such a program is called an assembler and in practice it may include a number of other functions. Assembler programs are the first step towards making computer systems easier to use. Assemblers provide a number of user conveniences but apart from macro-instructions they still retain an essential one to one correspondence between an instruction as written and a hardware instruction. The systems we will be discussing in the next chapter carry these concepts a great deal further and bring the language of the machine much closer to that of the user or application area. The problem is then how to convert programs

written in higher level languages into equivalent programs in the hardware instruction set of the machine.

14.4 Exercises

In the following problems, assign constants and storage as necessary.

1 Test the unknown integer in location 600. If it is odd and positive go to 300. If it is even and positive go to 300. If it is even and negative go to 400.

2 Given 50 numbers stored sequentially starting at 200, move a copy of them to sequential memory location starting at 500, and in so doing count how many of them are non-zero.

3 If the magnitude of the word at 400 is equal to the magnitude of the word at 600 jump to 500, otherwise jump to 700.

4 Use a repeat loop to count the number of words lying between 200 and 299 inclusive that have a positive sign.

5 Write a program which reads an alphabetic card into an input buffer starting at 10. If columns 11 – 14 of the card contained "NAME" print your name on the typewriter and punch it on a card.

14.5 Suggested reading

1. "The Elements of Digital Computer Programming".

   E. D. Reilly and F. D. Federighi (Holden Day).

2. "Fundamental Algorithms".

   D. E. Knuth (Addison Wesley)

3. "Fortran IV Programming".

   R. S. Ledley (McGraw Hill).

Chapter 15

ELEMENTS OF SOFTWARE SYSTEMS

THERE IS ONLY ONE CONDITION in which we can imagine managers not needing subordinates, and masters not needing slaves. This condition would be that each (inanimate) instrument could do its own work, at the word of command or by intelligent anticipation, like the statues of Daedalus or the tripods made by Hephaestus, of which Homer relates that "Of their own motion they entered the conclave of Gods on Olympus" as if a shuttle should weave of itself, and a plectrum should do its own harp playing.

ARISTOTLE
The Politics, ca. 350 B.C.

Chapter 15

ELEMENTS OF SOFTWARE SYSTEMS

15.1 Software systems

The term software is generally used to describe the programs as distinct from the hardware associated with a computer system. In broad terms we can distinguish two classes of software although there is a great deal of over-lap between these classes. The first class is problem or application oriented and is designed to make it easy for the user to write programs for different application areas. In this class we include language translators or compilers and various application oriented systems. The second class is largely machine oriented, the primary concern being to ensure that the system resources are working as effectively as possible. We include in this class operating or supervisor systems. The need for this latter class of software arises from the speed and complexity of modern machines which make human control quite impractical, and from the use of computers in control and monitor situations.

Application software

The first programming systems were sub-routine libraries and these remained as the main aid to users until the early fifties. However, the difficulty of using machine languages provided an incentive towards the development of more advanced programming languages. The early languages were interpreters, the action of which is to interpret one statement at a time of a more advanced language. These systems gave way to compiler techniques in the middle to late fifties but are now making a strong comeback as a result of the demand for interactive terminals and the development of microprogrammed computers.

Possibly the most important and impressive development in the fifties was the first FORTRAN compiler. This required some 25 man years of effort and was first released in 1957. Even so it produced problems and did not become readily accepted for several years. Nowadays such languages are taken for granted.

Another significant development was that of ALGOL. This resulted from the activities of a committee (mainly of European interests) who were interested in the design of algebraic compilers and decided to try and reach agreement on a common language which would be implemented on various machines. This would be a computer-independent problem-oriented language which was a concept with many obvious advantages. The group was widened and supported by others. ALGOL is now implemented on many machines and is widely accepted as both a programming and documentation language.

Comparable developments were soon taking place in the field of commercial data processing. The accepted standard language is COBOL. The responsibility for definition and acceptance of COBOL was due to a U.S. Government committee which very simply enforced standardisation by making it a requirement for all systems supplied to the U.S. Government.

Languages such as FORTRAN, ALGOL and COBOL are generally classified as procedure-oriented languages: programming languages which describe a process to be performed in terms of the procedures to be employed and the data to be processed. Such languages must be able to specify in detail all the mathematical, and logical procedures the user wishes to perform and should do so in a form as close as possible to the accepted practice of the relevant application area. Thus one of the features of scientific languages (ALGOL, FORTRAN) is the close correspondence between the method of writing mathematical expressions for evaluation and that of normal algebra. In addition loops and control statements are readily interpreted from their close correspondence to English. The user must also describe in detail the form of the data he wishes to be processed. This is done by naming them, generally using mnemonics and if necessary defining their dimensions magnitudes, forms, format, bounds etc. For example, if numbers are used he must state whether they are real, complex, floating point, integer etc. and where arrays and matrices are involved their dimensions and components must be listed. Likewise when data files are used their contents, format and size must be precisely determined. One of the interesting developments in recent years has been in languages for non-numeric data; for example, data arranged in strings (SNOBOL) or in interconnected lists (LISP). These needs arise from the application of computers to non-arithmetic areas in which these forms of data representation occur.

A second group of programming languages are *problem oriented:* these languages describe a problem to be solved in terms of the desired results or of problem statements. A problem-oriented language so defined is at a higher level of abstraction than a procedure oriented language. Its use implies that the programming language processor has the capability of decoding the source language into the equivalent of a procedure oriented language during the process of producing a target language suitable for machine execution. Such languages are a more recent development and will undoubtedly play a very significant role in the future, particularly in engineering design. Probably the major achievement to data is ICES (Integrated Civil Engineering System) which was developed at MIT. This is a comprehensive system simplifying computer use in many areas of civil engineering practice. Several systems are also available for circuit analysis in electrical engineering of which ECAP (Electronic Circuit Analysis Program) is probably the best known. Such systems enable an engineer to easily investigate a much wider range of alternatives than would be otherwise possible and so achieve a much higher degree of design optimisation.

It is not the intention of this course to acquaint the user in any detail with the use of either procedure or problem

oriented languages. The assumption is that readers have had some experience in the use of such languages and it is the aim of this chapter to give some insight into the theory and practice of their design and implementation. There are many excellent texts covering the use of such languages and the reader who is unfamiliar with them is referred to these sources. In particular a number of very good self instruction manuals are now available.

It is convenient at this point to introduce some terminology and definitions.

Source language: the language being described. In general this refers to the programming language under consideration.

Metalanguage: the describing language; that is the language used to describe the source language. This language also of course has its conventions and restrictive features.

Target language: the language produced from the source language by the programming language processor or by the decoding process.

Formal description of programming languages

To avoid ambiguity, encourage precision, and to identify clearly the details of structure, various methods of formal description are now employed. Such methods have the aim of being unambiguous and complete in the sense that they enable no alternative interpretations and describe all of the allowed constructs of the language. They are also compact and easy to understand so that they provide a convenient reference.

The natural starting point is to firstly describe the allowed symbols of the language and then the various rules for construction of the allowed "words" from these symbols. The structure of the equivalent of "clauses", "sentences", "paragraphs" and the like then follow. The comparison may be continued – in natural languages sentences may be classified as: declarative, imperative, interrogative and exclamatory. Statements which are the equivalent of sentences in programming languages are also classified as: declaration statements, assignment statements, procedural etc.

We refer to the definition of the structure of a language by the term syntax. One form for presentation of the syntax of a programming language which has become quite popular since it was first used by Backus et al in describing ALGOL has become known as Backus normal form or BNF. The principal features are as follows: –

(a) The symbol ::= is a separation symbol separating the right and left sides of a definition.

(b) The symbol | separates alternatives

(c) The brackets < > are used to enclose non-terminal symbols.

(d) Terminal symbols stand by themselves without brackets.

(e) A definition has the form

<entity being defined> ::= <definition 1> | <definition 2>
<definition m>

Definitions are commonly recursive in that an entity appears in both left hand and right hand sides. This is not ambiguous provided the definition includes a non-recursive construct as a prior alternative. We will illustrate the use of BNF by defining some selections from FORTRAN. A complete definition of FORTRAN is quite extensive but the following should illustrate the principles: –

<digit> ::= 0|1|2 |9

<letter>::= A|B|C |Z

<arithmetic operator> ::= +|-|*|/|**

sign ::= +|-

We introduce here the notation $\overset{n}{::=}$ which means that this definition can only be used n times for a construct to be valid.

<unsigned integer> $\overset{i}{::=}$ <digit> | <unsigned integer><digit>

<integer> ::= <unsigned integer> | <sign> <unsigned integer>

<identifier> $\overset{6}{::=}$ <letter> | <identifier> <letter> | <identifier> <digit>

<implicit integer letter> ::= I | J | K | L | M | N

<implicit integer name> $\overset{6}{::=}$<implicit integer letter> | <implicit int. name>

<letter> | <implicit integer name> <digit>

Reference (1) includes a formal description of FORTRAN IV.

Implementation of language processes

The object of a language processor is to translate the source program into an equivalent program in machine code. We distinguish two different strategies although in practice both may be employed – interpreters and compilers.

An interpreter carries out statement by statement translation – each source statement is translated and then immediately executed. It is somewhat analogous to sentence by sentence translation of a spoken language. Interpreters are relatively simple to write but have the disadvantage that interpretation must take place each time the program is run and this is expensive in computer time. For this reason they fell into disrepute but are now becoming popular again.

A compiler translates the entire program into a target language before execution is entered. Thus a program which may be run repetitively can be held in translated form and will then execute at much higher speed than the corresponding interpreter. However, it is not unusual to define a pseudo or artificial language as the target language for the compiler and for this target language to be executed via an interpreter.

Implementation of a translator involves two main phases – recognition and generation. Recognition involves determination of the structure of the statement being analysed as a result of which the equivalent code in the target language can be generated. There are several techniques which are commonly employed and therefore worthy of mention.

Polish notation (Lukasiewicz) is a method of writing arithmetic expressions which is easy to translate to machine code. It avoids the use of brackets by placing the operator always immediately right of the corresponding operand pair. The rule is to scan from left to right – whenever an operator is encountered, apply it to the two preceding operands and call the result a single operand. The following examples illustrate the principles: –

(X – Y) * (A – B) is written as XY – AB – *

(b * c) + (e^f + d) is written as bc * ef ↑* t where ↑ denotes exponentiation.

It is a relatively simple procedure to translate from Polish to equivalent machine code. Translation from conventional expressions to Polish is readily accomplished with simple algorithms which re-arrange symbols using stacks for storage. A stack is an addressless form of store based on a last in first out (LIFO) method of retrieval.

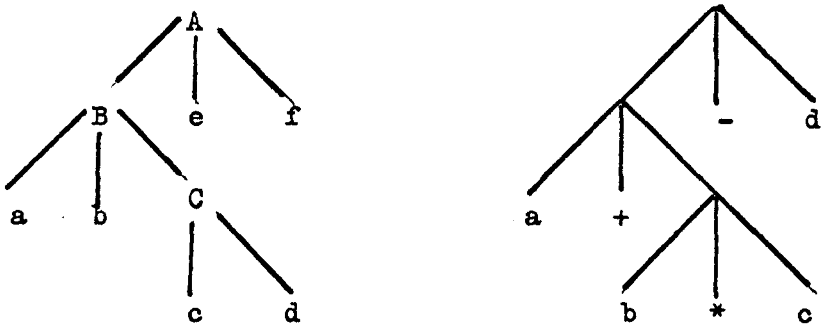
A tree is also a popular form of representing the structure of a statement (Fig. 15.1A)

The nodes of the structure are ABC and terminating symbols a b c d e f. The tree structure also leads to simple determination of the order of execution. For example, in Fig. 15.1a, execution proceeds from the bottom and the structure is that of

(a + (b * c)) - d

Two major steps are involved in translation – recognition and generation (Fig. 15.1b). The object of recognition is to determine the structure of the source language statements. In general this involves at least two phases; scanning for symbol recognition, and analysis to determine syntactic structure. Violations and ambiguities in source statements will also be detected and lead to diognostic messages. The output of the recognition phase is a representation of the structure which is used to synthesise a sequence of macro-instructions. In general this sequence needs some further attention to produce efficient code after which the equivalent program is emitted in the target language. The program is then ready for execution and may be used as often as needed without further translation.

We distinguish two methods of realising the above. The first method is the older and conventional one in which a number of programs are coded to achieve the above objectives in accordance with the language specification. Thus there is a



A,B,C are nodes a,b,c,d,e,f are terminal symbols

Representation of (a + (b * c)) - d

FIG. 15.1a Structural Relationships may be Represented as Trees.

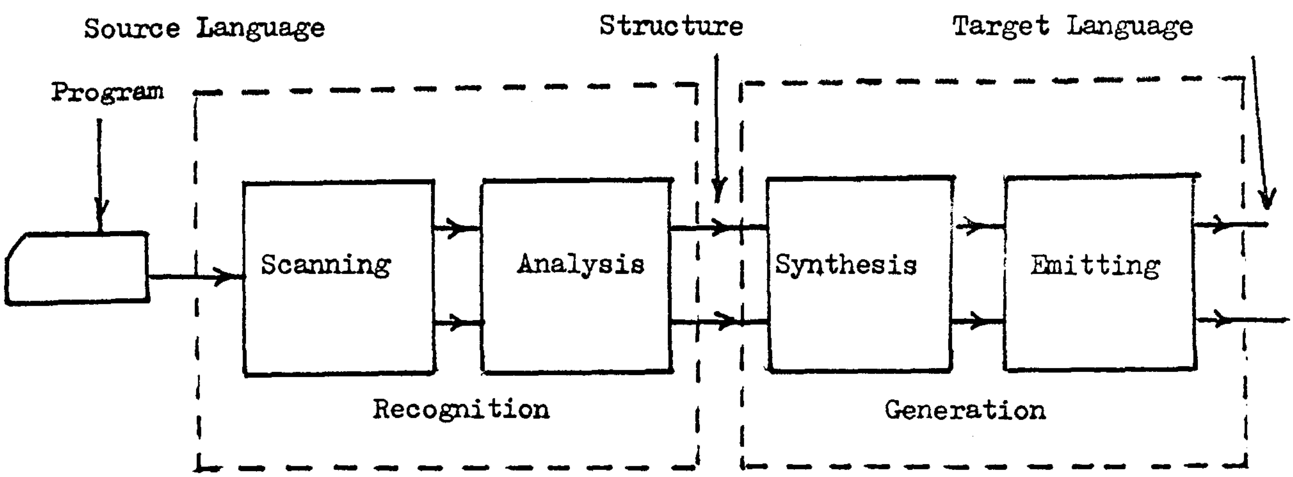


FIG. 15.1b Translation from Source to Target Language by Computer Program

lack of generalised strategy and alterations are difficult to make. The second method uses tables to encode the syntactic specifications and uses an analyser algorithm to perform analysis on the source program using the tables as reference. This more formal method allows changes to be made without great difficulty and is generally preferred.

There is much that still remains to be done in the area of language processors and it is an active area of research. Some of the current research topics are compiler compilers, extensible languages, graphical languages, and advanced problem oriented systems. At the same time hardware designers are providing more effective machine functions to reduce the burden of language processor design.

15.2 Machine software

The previous discussion has been concerned with software designed to make computer systems easier to use. We now concern ourselves with software systems where the principle objective is making the most effective use of system resources. At

the same time such systems provice the user with services which allow storage and management of data sets and enable separately compiled programs to be linked together. The term operating systems is generally used to describe such software, although supervisor systems or monitor systems are sometimes used.

The need for operating systems became clearly evident with the development of transistor-core computers in the late fifties. The total problem solution times in these and later systems may be only a fraction of a second. During the same period progress in input-output equipment was relatively slight so that the combination of slow input-output devices and very fast processing units produced problems in efficient system utilisation. The simplest solution to input-output mismatch is to use an off-line computer to transcribe source programs from relatively slow media such as card readers to fast media such as magnetic tape. However, this is only a partial as well as a somewhat expensive solution. The ultimate solution sought was to have the computer supervise its own operation and to request attention from a human operator only when needed to make a decision or to load or unload a peripheral unit. At the same time this concept created a natural desire for more sophisticated systems. In particular systems were designed which allowed the processing unit to "time slice" operation amongst a large number of jobs which may originate from multiple input terminals.

It is not unreasonable to draw an analogy between a computer system and a production factory. Perhaps the most outstanding difference is that apart from input and output a computer does not deal with physical quantities but rather with the manipulation of abstract quantities, which could however represent models of physical situations. In other respects there is a great deal of similarity. A computer system consists of an assembly of production resources, some hardware, such as input-output units, storage and processing, and some software, such as language processors. The work load consists of one or more input job streams, each of which can have different demands on resources and a number of special requirements such as high job priority. The problem is to design an operating system which can satisfy all of these needs. As yet there are few formal approaches to the design of such systems. However, certain of the concepts involved and facilities provided are of interest and will be discussed.

One of the most apparent problems is to ensure protection of the supervisor system against mutilation due to errors in user programs. It is usual to provide some hardware features which guard against this situation. A typical one is to define two states for the machine, a supervisor state and a problem state. In this strategy some machine functions are only available when in the supervisor state and proper choice of these can give the requisite protection.

A typical operating system consists of a number of major components for which a typical sub-division would be: –

Job scheduling: A set of programs are provided to schedule the job stream into and out of the system.

Input-output control: It is normal to refer all requests for input-output to the operating system which can then assign units to jobs and initiate the requisite routines. Thus the system is able to maintain status information on all peripheral units which call the system by means of interrupt signals whenever there is a change of status – e.g. completion of a transfer operation.

Link edit: This enables sub-programs which may have been independently written or form part of the library to be interconnected into a master program.

Utility routines: These provide services such as console messages, monitor and dump facilities.

System monitor: This section is entered for all abnormal conditions and initiates the appropriate response.

Data management: This provides a variety of services to enable orderly organisation and retrieval of data files.

The following diagrams illustrate some of the activities of an operating system. Fig. 15.2a shows a simple system for sequential scheduling of jobs through a system. The input job stream may be entered on cards, tape or disk and each job is first examined by a reader-interpreter which notifies the master scheduler. As soon as the scheduler is notified of completion of the present job it can then initiate the next valid job available.

Fig. 15.2b shows a more advanced scheme. In this scheme input and output are made separate tasks. This enables a queue of input work to be built up on a faster input device by the reader interpreter. Jobs may be taken from this queue at much higher speed and in priority sequence. Likewise, output is queued on a faster medium and independently extracted by an output writer. This system is capable of smoothing the individual mismatches of input-output to computing and can increase throughput very greatly.

Fig. 15.2c shows the independent phases involved in the passage of a job submitted in a higher level language. The source program may request any of a number of available language processors and may be entered in a variety of media – cards, tape, disk etc. The language processor then acts on the source program to produce an *object* module. This is still not ready for execution and requires to be processed by the linkage editor which joins the various components together including library routines to form a *load* module. This is the form needed for loading and subsequent execution.

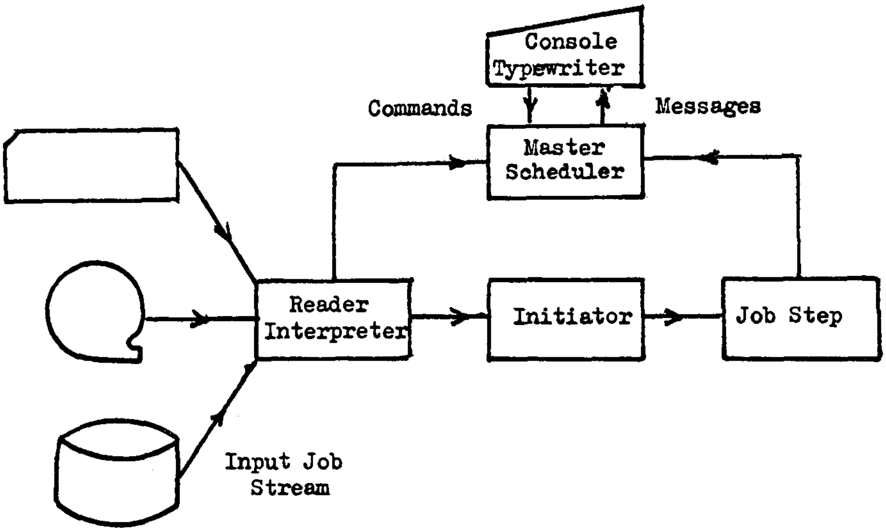


FIG. 15.2a Organisation of a Simple Sequential Scheduler

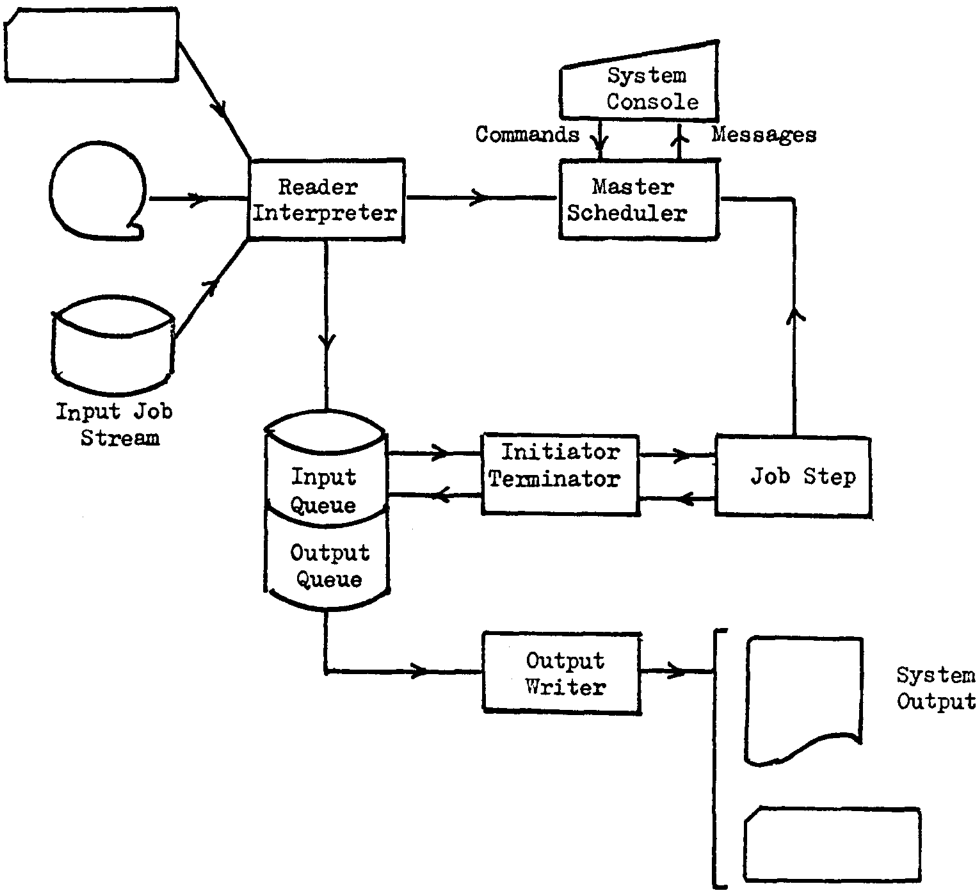


FIG. 15.2b Organisation of a Buffered Priority Scheduling System

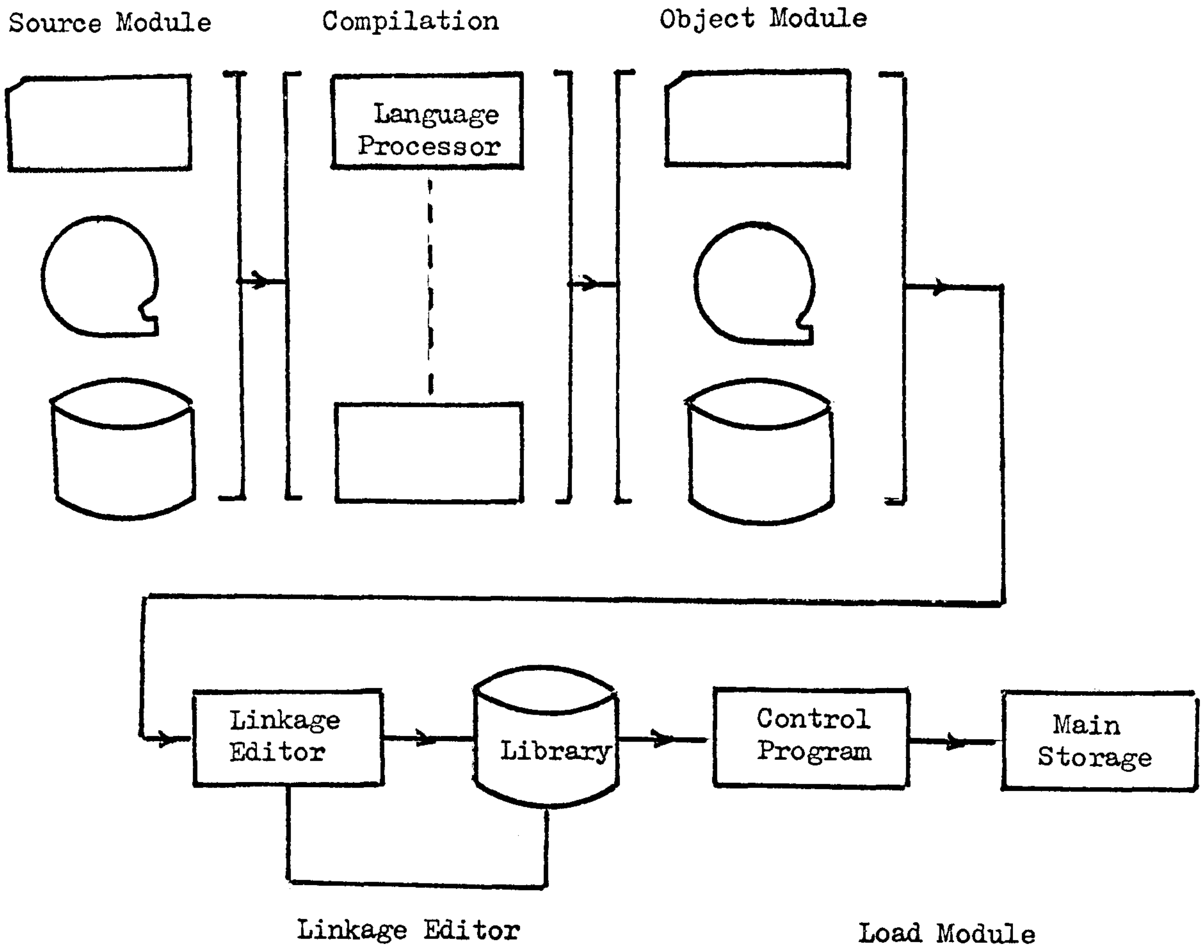


FIG. 15.2c Illustration of Phases Involved in Processing a Job

Nowadays there is a great deal of interest in multi-access computing systems which can simultaneously service a large number of user terminals. Such systems make heavy demands on operating systems in order to provide adequate protection amongst users and efficient use of system resources.

As systems become larger, faster, and more complex the corresponding demands on operating systems will increase. Research is needed to develop better formal methods of describing and implementing operating systems.

15.3 Exercises

1 Using the definitions given in this chapter –

(a) Which of the following are allowable identifiers –

(1) A

(2) 3A

(3) ABCDEFG

(4) A.CD

(5) -ABC

(b) Which of the following are valid integer names –

(1) I3

(2) IBC

(3) KX

(4) N.X3

2 Convert the following expressions to Polish notation –

(a) $ax^2 + bx + c$

(b) $e^f (a + b) \quad (c - d)$

(c) $(ax + b) \quad (cx^2 - d)$

3 Can you suggest some good areas for the development of application oriented computer software.

4 What applications can you suggest for on-line computer terminals in education.

15.4 Suggested Reading

(1) "Fortran IV Programming".

R. S. Ledley (McGraw Hill).

(2) "Scientific American September 1966". This is a special computer issue.

(3) "Operating System 360 – Concepts and Facilities". I.B.M.

Try and obtain some manuals on computer software and read them carefully.

Chapter 16

CONTROL OF LARGE SYSTEMS

16.1 **Introduction**

In this chapter we will not only discuss the technical aspects of the control of large systems, but also will see how the engineer approaches such problems, analyses them, devises methods of control, and designs devices for implementing the control.

In discussing this subject we will use the three languages of engineering; firstly the language of ideas and knowledge using the medium of the written word; secondly, the language of patterns, using diagrams, and thirdly the language of logic using the symbols and rules of mathematics. By using each of these three languages where appropriate we will be communicating as one engineer would to another.

Let us begin by considering some of the key terms used in systems engineering, and by examining several of the basic notions employed. In this we will include a discussion of the concept of state and of state-space, and the idea of motion of a system in terms of movement of a point in state-space. We will see how this motion may, to a limited extent, be controlled by deliberate control action, and how these dynamical properties and relationships can be described explicitly in terms of mathematical equations, particularly the so-called state equations.

Particular mention will also be made of control decision sequences, and attention drawn to the wide variety of sequences ranging from the most elementary sequence comprising a single control decision, to the most complex involving an infinite sequence (or a continuous function in time).

We will then draw attention to the system performance criterion, by which the effectiveness of any particular control decision sequence can be measured in quantitative terms.

To see more clearly how these terms and ideas apply in practical engineering, we shall consider several particular systems. We shall begin with a small system capable of precise description in the terms of system engineering, namely a satellite attitude positioning system. This will be followed by the slightly more complicated motor-vehicle displacement positioning system, where the dynamical properties are less precisely known and the optimum control decision sequence more difficult to determine. We will then draw attention to launching and positioning systems of a synchronous satellite, point out the technical problems involved in estimating the state (or orbit) of the satellite and indicate how the uncertainties here influence the form of control decision sequences employed. We then begin to enter the realm of large systems by considering a simplified electrical power generating, transmitting and utilizing system, in which the object of the control decision sequences is to satisfy the load demand in the most economical way. In these first four systems we will observe that they may be discussed and analysed quantitatively in the abstract terms of diagrams and of mathematics.

We then go on to consider several systems which are far more complex in their structure and in the interrelationship between variables, and whose dynamics can at the best only be described in mathematical terms if the very crudest of approximations are used. Despite their complexity, they are however systems amenable to control by humans in a reasonably stable if not optimum way. In this group we will focus upon road-traffic systems which are generally not very well suited to abstract modelling in mathematical terms, but which to some extent can be simulated by programmed sequences in digital computers. We will see that a sound logical basis for the design of an optimum controller is exceedingly difficult to find because of the complexity of the process being controlled. An even more complex system is that encountered when control decisions are to be made on the placement of breakwaters in harbour installations. So complex is the hydraulic system in such instances that abstract modelling in mathematical terms or by digital computer simulation even, is well beyond reach. To bypass this obstacle, engineers often build scale models of harbours (a physical as opposed to an abstract model) and learn by trial and error experimentation how the sand moves about the harbour. From this knowledge a single control decision is made on the size and location of a breakwater which will keep the shipping channel clear, and do so with the least cost. Finally, we take a brief look at the economic system.

As we proceed from the simple to the more complex systems, we will notice that whereas the systems themselves become increasingly difficult to describe in abstract terms and their performance criteria become harder to define, and the best decision sequences which it is possible to predict, necessarily become extremely simple. Whereas the optimum control sequence for satellite attitude positioning system may be precisely defined for each point in time over the control period, the corresponding control crude, comprising for example a single control decision as to the fund allocation for the next year.

16.2 **Definition of Terms and Equations**

State

The use of the term *state* in systems engineering closely follows the general usage of the term in its less technical sense. In engineering terms it means a set of values for the variables of the system, which uniquely define the condition of the system. For instance, a numerical statement on the altitude, position, speed and direction of flight of an aircraft might be a sufficient description of the state of the flying aircraft for certain purposes.

Generally it is found from observations upon a particular system of interest that there is a set of variables which are significant, and that a quantitative statement on these variables is fuficient adequately to define the state of that particular system.

Clearly, in all dynamic systems the state is continually changing due to either forces within the system or control forces impressed upon the system from outside.

State-Space

It is a useful abstraction to consider a *state-space* which has one co-ordinate for each state variable. Although such a space having one, two or three dimensions is easily visualised in the mind, there is no reason why the abstraction of higher order state-spaces cannot be accepted. If one prefers not to have a geometrical interpretation to the notion of state-space, then it is equally valid to use a state-vector having as many components as there are states of the system.

The geometrical interpretation is however convenient, and no complications in interpretation arise from considering fourth and higher order spaces as other than extensions of the notion of three-dimensional space of the physical world in which we live.

Using the geometrical approach, it follows that each point in the state-space uniquely defines (through its co-ordinates) the state of the system.

Motion of the System

As time passes and the state of the system changes, so the representative point in the state-space moves, and in so doing traces out a path in the state-space. This path is conveniently referred to as the trajectory of system in the state-space. It is convenient to speak of the motion of the system, and to associate this with the movement of the point in state-space as time passes and the state continuously changes.

Also the tangent to this line at various points gives the direction of *motion of the system.* So in speaking of the motion of a system we are essentially talking about the motion of the representative point in the state-space.

Control

As will be discussed in more detail later, the motion of a system depends both upon the state of the system and the control applied to the system, all considered at a stated point in time. Thus, although at a given point in time it is not possible instantaneously to change the state, it is possible to choose the control and thus to choose to some extent the direction of motion of the system.

Thus in general we note the very important observation that the motion of a system at each point in time depends upon the state of the system at that time and upon the control applied at that time. Hence through suitable choice of the control, the motion of the system may be determined.

State Equations

If we consider as an example a mass M moving in response to a force u, it is known from Newton's law of motion that the acceleration (rate of change of velocity) is given by

$$M\frac{d^2x}{dt^2} = u$$

where x = displacement of the mass

$\frac{dx}{dt}$ = velocity

$\frac{d^2x}{dt^2}$ = acceleration.

It is noted that this is in the form of a second-order differential equation, and it would be expected that, had we chosen a more complex physical system, the differential equation would probably have been of still higher order. Such differential equations are somewhat awkward to solve, and it is found in practice that there are advantages in changing the form of these equations into a set of first order equations. For instance, if we denote

$$x_1 = x = \text{displacement}$$

and

$$x_2 = \frac{dx}{dt} = \text{velocity}$$

and then it follows that

$$\frac{d}{dt} x_1 = \frac{dx}{dt} = x_2$$

and

$$\frac{d}{dt} x_2 = \frac{d^2x}{dt^2} = \frac{1}{M} u.$$

This may be written in the slightly more compact form

$$x'_1 = x_2$$
$$x'_2 = \frac{1}{M} u$$

where $x'_1 = \frac{d}{dt} x_1$ and $x'_2 = \frac{d}{dt} x_2$.

If we use vector and matrix notation which is briefly referred to in Chapter 9, we may write these equations in an even more compact form as

$$x' = Ax + Bu \qquad \text{...16.1}$$

or

$$\begin{bmatrix} x'_1 \\ x'_2 \end{bmatrix} = \begin{bmatrix} 0 & 1 \\ 0 & 0 \end{bmatrix} \begin{bmatrix} x_1 \\ x_2 \end{bmatrix} + \begin{bmatrix} 0 & 0 \\ \frac{1}{M} & 0 \end{bmatrix} \begin{bmatrix} u \\ 0 \end{bmatrix}$$

where

$x = \begin{bmatrix} x_1 \\ x_2 \end{bmatrix}$ = state vector

$x = \begin{bmatrix} x'_1 \\ x'_2 \end{bmatrix}$ = velocity vector

$u = \begin{bmatrix} u \\ 0 \end{bmatrix}$ = control vector which in this case has only one non-zero value.

$A = \begin{bmatrix} 0 & 1 \\ 0 & 0 \end{bmatrix}$ = constant coefficient state matrix.

$B = \begin{bmatrix} 0 & 0 \\ \frac{1}{M} & 0 \end{bmatrix}$ = constant coefficient control matrix.

The advantage of transforming one or more high order differential equations into a set of first order differential equations, and of using the vector notation is that a very wide class of dynamical systems may be described in the standard form of the so-called vector differential equation. The properties of these equations are well understood and standard methods of numerical integration by digital computers exist. In its general form the vector differential may be written

$$x' = f(x,u) \qquad \text{...16.2}$$

where

$$x = \begin{bmatrix} x_1 \\ x_2 \\ \vdots \\ x_n \end{bmatrix} \text{, the state vector}$$

$$u = \begin{bmatrix} u_1 \\ u_2 \\ \vdots \\ u_m \end{bmatrix} \text{, the control vector}$$

$$f(x, u) = \begin{bmatrix} f_1(x, u) \\ f_2(x, u) \\ \vdots \\ f_n(x, u) \end{bmatrix} \text{, the system vector function}$$

and

$$x' = \frac{d}{dt} x$$

For many purposes the particular canonical form associated with the so-called class of linear systems is simple and useful. For linear systems 16.1 becomes,

$$x' = Ax + Bu \qquad \text{. . . 16.2}$$

where A and B are constant matrices.

For linear systems we see very clearly that the motion of the system x' depends upon the state of the system x and the control u. We can see that in the absence of any control the system will undergo continuous movement, and that by the inclusion of control forces, this movement may be modified.

The state equations, whether in their general form in 16.1 or the more special form in 16.2 define in mathematical terms the dynamic properties of the system.

Control Decision Sequences

It is useful to consider controls as being a set of *decision sequences.* Sometimes controls are continuously applied to a system, and result from a continuous set of decisions. For instance in driving a car, one is continuously steering the car by continuously making decisions about the position of the steering wheel. At other times one is concerned to make a single decision or possibly a finite sequence of decisions. In driving a car there is a single decision required to switch on the engine, although after the car has started there is a continuous sequence of decisions required to regulate the engine speed to suit the driver's demands.

In the control of large systems, often possessing many controls, the task is to choose the control decision sequences which will result in the system operating in the best possible manner. It becomes apparent that for very simple and precisely defined systems, such optimum controls may themselves be also precisely defined. With larger and less well defined systems the optimum controls which one can determine are far less precise and for the most complex systems are frequently of the most elementary and crude form.

Performance Criterion

If it is accepted that the complete knowledge of particular systems is contained in the state equations for simple systems such as a satellite launching system, or in a digital computer simulation program in the case of a traffic flow system, or

in a physical model in the case of a harbour silting system, then it is implied that the motion of these systems can be modified to some extent by the choice and implementation of suitable control decision sequences. For quantitative measurement of the effectiveness of such decisions, clearly defined *performance criteria* are required.

It becomes apparent after a little thought that the form of a performance criterion is quite arbitrary, and depends entirely upon the objective the engineer wishes his system to achieve. In one case the requirement might be to drive the car between two points with the minimum of fuel consumption, where as in another situation for instance it might be required to drive the car between two points in minimum time. Whereas in the first case the criterion was the quantity of fuel consumed, and in the second it was the length of time elapsed. Clearly, the optimum control decision sequences in the two cases would be quite different, even though the physical system (motor vehicle and road) is identical.

Large System Control

We see that the effectiveness of the controls are measured in quantitative terms by evaluating arbitrarily chosen performance criteria. To determine which of a wide (often infinite) range of control decision sequences will be optimum in the sense of maximising (or minimising) the chosen performance criterion, it is necessary to have some means of predicting the relative effects of different control decision sequences. For simple systems whose dynamics can be adequately described by state equations, it is possible to use these as an abstract model from which the optimum control decision sequence can be obtained by suitable numerical experimentation. For larger systems of greater complexity, various approximations are required if suitable abstract mathematical models are to be obtained. For still larger systems, analogue models such as pilot plants may be built for predicting response. Ultimately for the very largest systems such as the economic system, one is forced into experimenting with the actual system, so that by continuously disturbing the system with small control perturbations, and observing the effects, one can build up some cause-and-effect information on which to make future predictions of the effect of certain control decision sequences.

16.3 Satellite Attitude Positioning System

When it is required that a satellite axis be turned for alignment with the earth, gas jets may be used to exert a torque and rotate the satellite. The arrangement of the jets is shown diagramatically in Figure 16.1.

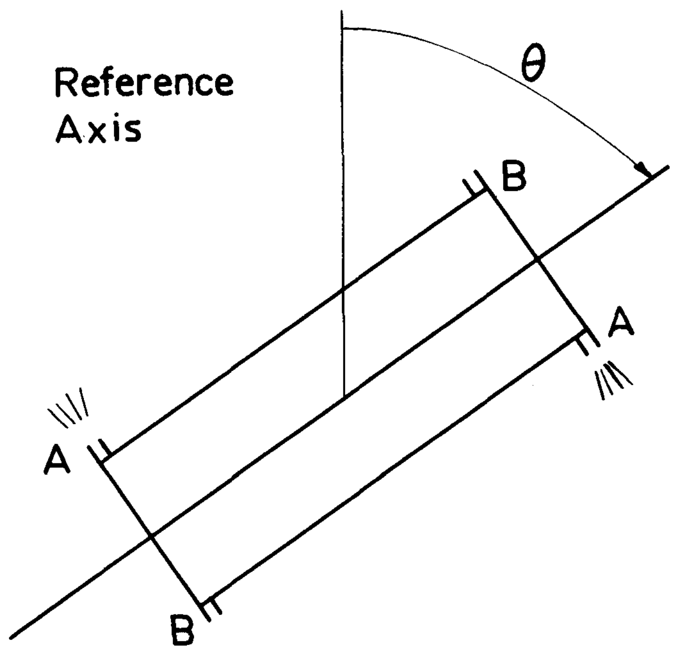


FIGURE 16.1

SATELLITE ATTITUDE POSITIONING SYSTEM

The system is one having precisely known dynamical characteristics. Its mathematical model may be written in terms of statespace equations (assuming for example unit inertia),

$$x_1'(t) = x_2(t), \qquad \ldots 16.3$$

$$x_2'(t) = u(t) \qquad \ldots 16.4$$

where

$$x_1 = \text{angular displacement} = \theta$$

$$x_2 = \text{angular velocity} = \frac{d}{dt}\theta(t)$$

$$u = \text{torque}$$

If it is supposed that the gas cylinder maintains constant pressure, then the torque is proportional to mass flow rate. For practical reasons, the maximum gas flow rate (and hence the torque) is limited. Suppose for the sake of example the torque is constrained such that

$$-1 \leqslant u \leqslant +1.$$

At the beginning ($t = 0$) of the control interval, it is supposed that the satellite is at rest with a unit displacement. That is

$$x_1(0) = 1$$

$$x_2(0) = 0$$

It is required to control the flow of gas so that at the end of a control period ($t = T$), the displacement and velocity be zero. That is

$$x_1(T) = 0$$

$$x_2(T) = 0$$

After a little thought it becomes apparent that the larger the control interval $[0, T]$, the less gas need be consumed. Thus assuming it essential that the least amount of gas be consumed, some upper limit must be placed upon the duration T. Once this has been decided it remains to determine a suitable control decision sequence. This in turn requires the choice of a suitable performance criterion. In this example, since gas consumption is to be minimised, it follows that the criterion should be that the optimum control decision sequence is that which minimises the index

$$e = \int_0^T |u(t)|\, dt \qquad \ldots 16.5$$

If in this example we choose T = 4½ seconds, it may be shown that a control decision sequence

$$u_a(t) = \begin{cases} -1, & 0 < t < 0.25 \\ 0, & 0.25 < t < 4.00 \\ +1, & 4.00 < t < 4.25 \end{cases} \qquad \ldots 16.6$$

brings the satellite to rest with zero displacement in the chosen time. It may be shown that this control decision sequence is not unique in achieving the terminal state and that for instance the control

$$u_b(t) = \begin{cases} -0.222, & 0 < t < 2.125 \\ +0.222, & 2.125 < t < 4.25 \end{cases} \qquad \ldots 16.7$$

will achieve the same result in the same time.

Two common patterns used to describe the effect of the controls u_a and u_b from equations 16.6 and 16.7 respectively, are the graphs showing the control as a function of time in Figure 16.2 and state-space diagram showing the trajectory in state-space of the system in response to these two controls as shown in Figure 16.3. It is interesting to note that patterns of both these kinds are widely used in the study of the response of dynamical systems.

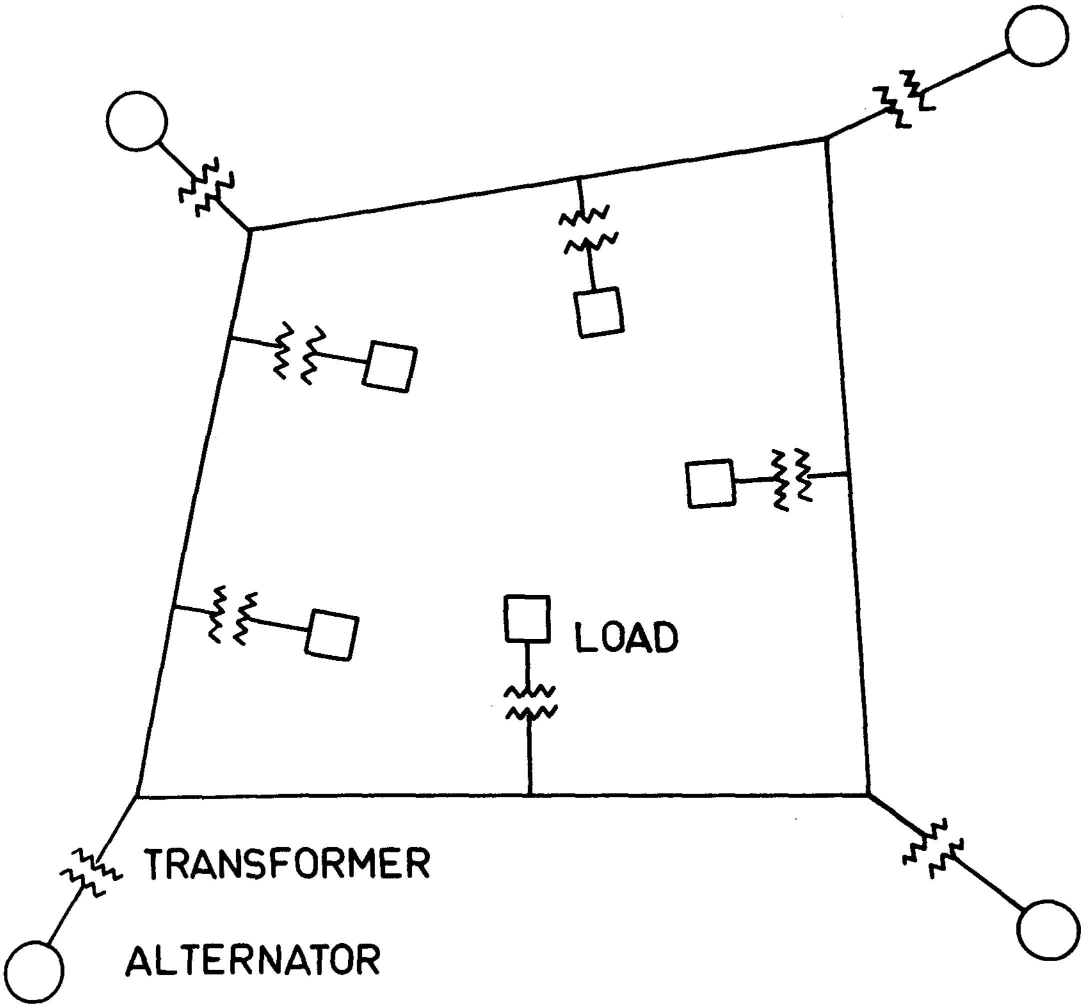


FIGURE 16.2

ELECTRIC POWER SYSTEM CONNECTION DIAGRAM

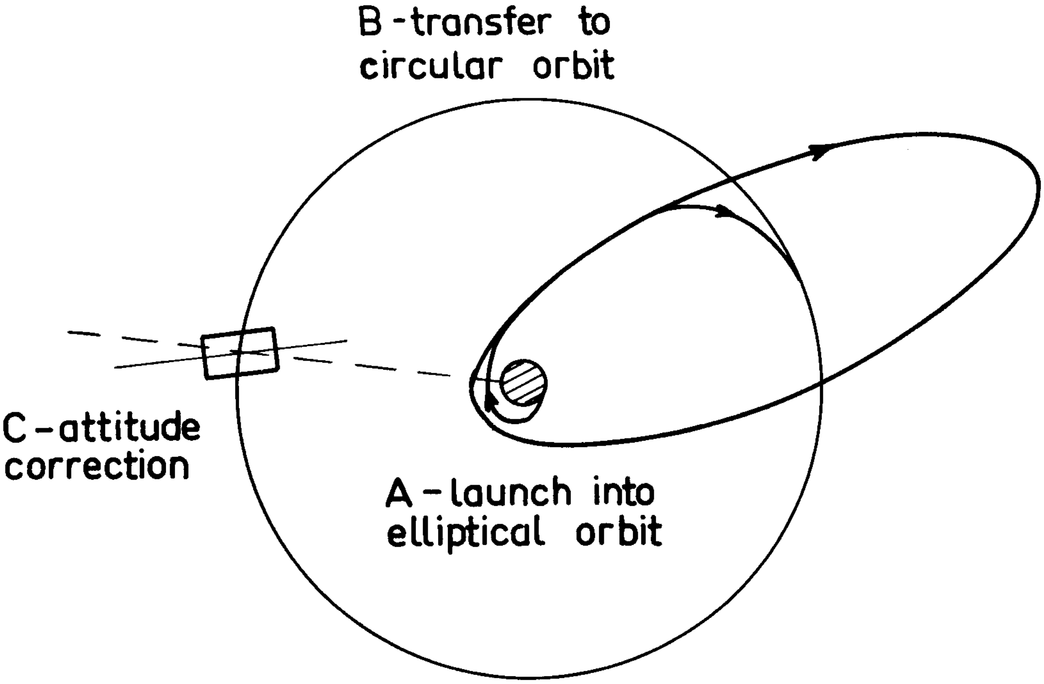


FIGURE 16.3
SATELLITE LAUNCH ORBITS

A number of important observations may be made about this example:

(a) The system is a very simple one and its state may be completely defined by only two state variables (x_1 = displacement and x_2 = velocity).

(b) The corresponding state-space being two-dimensional may be conveniently depicted on a sheet of paper using a pair of rectangular co-ordinates, to locate successive states and thus to describe a trajectory showing the motion of the system.

(c) Because of the simplicity of the system and the precision with which its response may be calculated, the effectiveness of the various alternative controls in reducing the value of the error index may be accurately compared.

(d) The dynamical characteristics of the system are accurately described, without approximation, by two simple equations 16.3 and 16.4

(e) The control decision sequences may be precisely described using if necessary, a very large number of control values, one for each of a correspondingly large number of points in time during the chosen control interval.

We will now proceed to consider systems of increasing complexity and we will observe that we quickly will lose the simple precision of the present example. We will, in the more complex systems, lose the ability to employ the valuable abstraction of mathematical models. Also, the fine detail of the control decision sequence, possible in the satellite attitude positioning system, will no longer be possible to achieve, and at the best a crude approximate single valued control sequence will be all that can be obtained with the present knowledge of the dynamics of the systems concerned and with the limitations of existing computers in their power to simulate these systems.

These difficulties throw up vast new challenges to engineers in the future, to find better ways of determining the dynamics of systems, better ways of forming abstract mathematical models of such systems, and better ways of using analogue and digital devices to simulate the performance of complex systems.

16.4 **Vehicle Positioning System**

Let us consider a motor vehicle moving along a roadway, starting from rest at the traffic lights at one intersection and coming to standstill at the lights at the next intersection. The driver can use the accelerator and the brake, and may wish to decide how best to use these in order to get from one set of lights to the other in say minimum time.

At first sight this is a comparatively simple systems engineering problem, but in closer inspection it will be seen that there are difficulties. The first difficulty is that of obtaining an accurate abstract mathematical model of the systems. If we are prepared to assume that the frictional force is proportional to the square of the speed (a very rough approximation), the dynamics may be described by the state equations,

$$x'_1(t) = x_2(t) \qquad \ldots 16.8$$

$$x'_2(t) = -x_2(t)^{\,2} + u(t) \qquad \ldots 16.9$$

and where the initial state is

$$x_1(0) = 0$$

$$x_2(0) = 0$$

and the final state is $x_1(T) = K$ (where K is the distance between intersections)

$$x_2(T) = 0$$

and where time T is unknown.

As a further gross approximation let the maximum possible force of acceleration be the constant u_c and the maximum possible force of braking be u_d.

For those who have access to a digital computer, and are interested in solving difficult problems, the determination of the control decision sequence which minimises the transit time between traffic lights will provide a hard challenge. (Use $K = 1, u_c = 1, u_d = -2$, say).

16.5 **Satellite Launch System**

We now turn to a system of greater complexity in which the dynamical characteristics are fairly well known but where disturbing forces of unknown magnitude are present. Because of these unknown disturbing forces, it is not possible to predict accurately the future state of the system, and so it becomes necessary to take steps to measure the state at future points in time, using an independent measuring system.

Although the satellite can be placed into an elliptical orbit during launch, the lack of precise knowledge of the dynamics of the launcher during flight through the atmosphere, and the presence of unknown disturbing forces, make it impossible to ensure that the satellite will achieve exactly the desired orbit. However the precise orbit must be determined to enable a subsequent transfer to be made to a circular orbit, and this is achieved by use of radio signals. By means of directional antennae on earth, it is possible to make numerous observations of the satellite's position as it moves in its elliptical orbit, and subsequently by statistical analysis of these data in electronic computers to obtain an increasingly accurate estimate of the orbit.

In the next phase of the operation in which the satellite is transferred to circular orbit, the dynamics of the system are precisely known and there are only negligible disturbances present. The forces required to cause the transfer are capable of precise calculation and exact implementation, so that provided the transfer commences when the satellite is at the correct point on the elliptical orbit, the circular orbit will be acquired with corresponding precision.

The final phase of attitude correction has been described above.

Although during the launch operation the number of states required to describe the condition of the system at each point in time are fairly numerous, and the number of state equations required to describe the dynamics of the system are equally numerous, the numbers are still quite manageable, and hence the use of abstract mathematical models and computer calculation of the trajectories in state-space is quite practicable with modern computers. As stressed above, the complete sequence of control decisions cannot be calculated completely in advance because of the presence of disturbing forces which cause the satellite to deviate in unpredictable ways from the planned orbits. It becomes necessary to introduce intermediate steps in which the elapse of time is required accurately to measure the resultant state of the system. This new information may then be used for calculating the next steps in the control decision sequence. This process of measurement of state and consequent adjustment of control is called feedback, and as will be seen below, the less precisely the system dynamics and disturbances

are known the greater is the need for feedback if desired objectives are to be achieved.

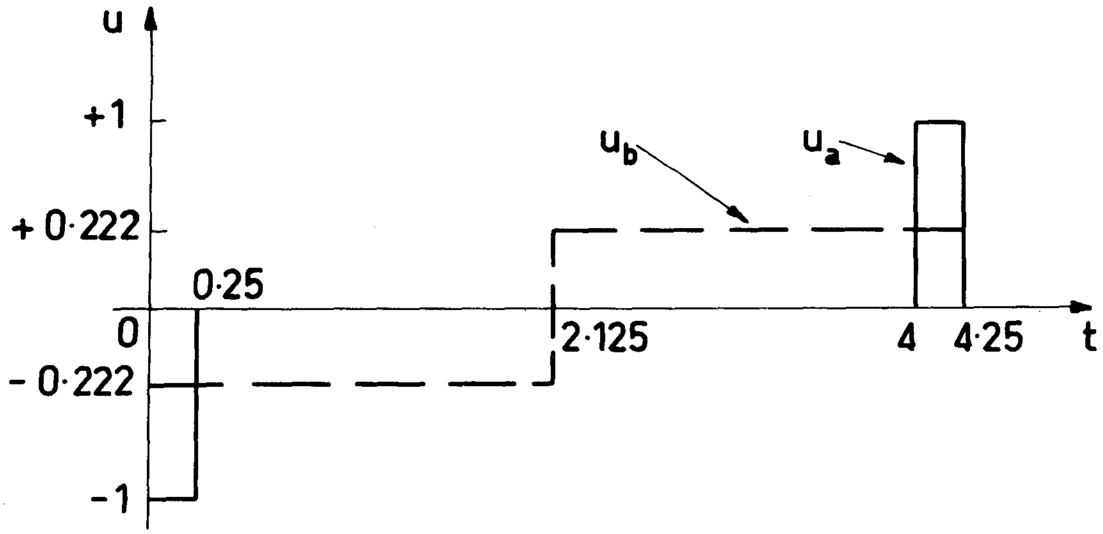


FIGURE 16.4

TWO POSSIBLE CONTROL DECISION SEQUENCES $u_a(t)$ FROM EQUATION (6) AND $u_b(t)$ FROM EQUATION (7)

16.6 Electric Power System

It is apparent that electric power systems contain very large numbers of components, and thus an exact description of the state of the system would require a prohibitively large number of state variables. Because of this it is usual to make a number of simplifying assumptions, to pick out a number of the more important variables and to use these as state variables. A correspondingly approximate description of the dynamical characteristics of the system is given by a set of simplified state equations.

Because of the approximations involved, the state equations forming an abstract mathematical model of the system, would not be used for accurate prediction of the future states of the system. Rather, it is necessary for extensive measurements to be made continuously upon the system to determine the state with sufficient accuracy.

Such system models are however sufficient for some control purposes, and may be used for instance when the criterion of performance is that of supplying the demand for electric power at theseveral load points, but to do so by generating the power in the most economical way. It is known that some generators in a system are less efficient than others, and that power is lost if the generating station is a large distance from the load. By continuous measurement of the state of the system, and by use of an approximate mathematical model of the system, it is possible to take into account the continuously changing load condition, and thus to establish a continuous sequence of control decisions for generation.

We see that electrical power systems may be controlled by using simplified mathematical models which enable approximate estimates of system responses to particular controls to be estimated. We note also that it is a system subjected to large disturbances as users change their demand for power in a largely unpredictable way. The performance criteria are however generally fairly simple and easily defined.

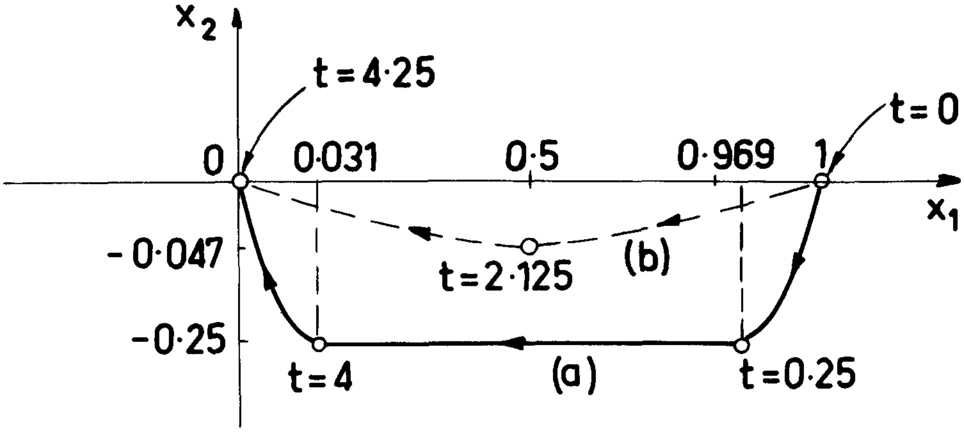


FIGURE 16.5

TWO TRAJECTORIES IN STATE SPACE CORRESPONDING TO u_a AND u_b FROM EQUATIONS (6) AND (7)

16.7 Vehicular Traffic System

A vehicular traffic system comprising a network of intersecting roads, and a flow of vehicles traversing the network, is one of considerable complexity. The states of the system are difficult to define, and the dynamical relationships are equally difficult to establish. It is even difficult, if not impossible at present, to establish a simplified mathematical model of the system.

The control decision sequence comprises the statement of the time at which the lights at each intersection change colour. Considering the large number of intersections which may be involved, and the frequent changing of the aspects of the signals which is required, we can see that a very large statement of control decisions is required.

Furthermore, the criterion to be used for stating the performance required of the system, for use in determining the optimum control decision sequence is not by any means obvious. One might decide that the total transit time for vehicles entering the system during a defined period should be minimised. However this might give very poor service to traffic in some streets and because of this a more comprehensive criterion might be needed for a practical situation.

It is important to bserve that despite the complexity of the system dynamics and the uncertainties about the appropriate performance criterion, vehicular traffic networks do function reasonably well. However it is unlikely that they function in the best possible way, and it is to be expected that more sophisticated controls would give better system performance.

16.8 **• Hydraulic Systems**

Some of the most complex systems to be controlled are found in harbour installations. Here we have large masses of water carrying silt and often depositing this in unwanted places. Very often special (and very expensive) breakwaters have to be built to deflect the flow of water and avoid the silting-up of important navigational channels. The control decision sequences involved are clearly of a single-shot nature, and there is no room for experimenting with the system itself. The decision as to where to place a breakwater and what size to make it must be correct first time.

The dynamics of such hydraulic systems are so complex that mathematical modelling and computer simulating (at present) are not possible. It has been found however that small-scale physical models may be used to considerable advantage, and that prediction of the transportation of sand by water current may be accurately estimated. Thus by systematic experimentation with the placing of breakwaters in the model, it becomes possible to decide the most economical and suitable site.

We note that in the control of this system there is very little place for the application of the feedback principle. The size and location must be correctly decided (from experiments with the physical model) in advance of the implementation of the decision to construct the breakwater. Some minor modifications and extensions might eventually be required but even these should be at least anticipated in the making of the main decision.

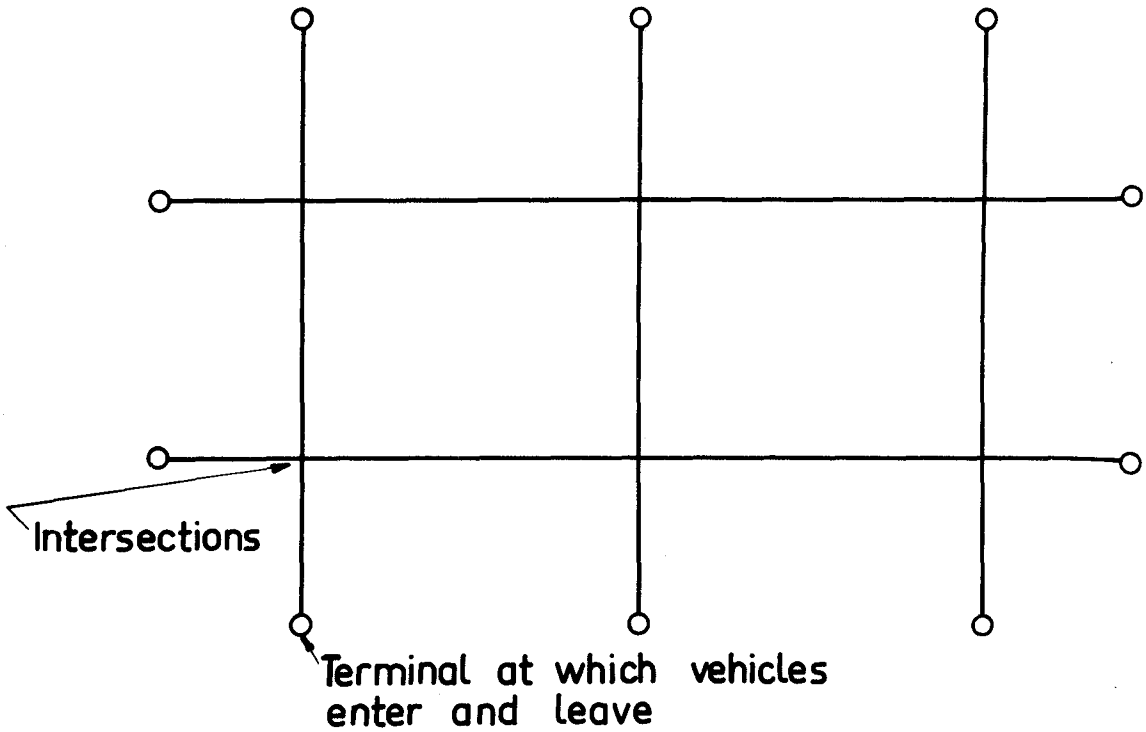


FIGURE 16.6
ROAD TRAFFIC NETWORK

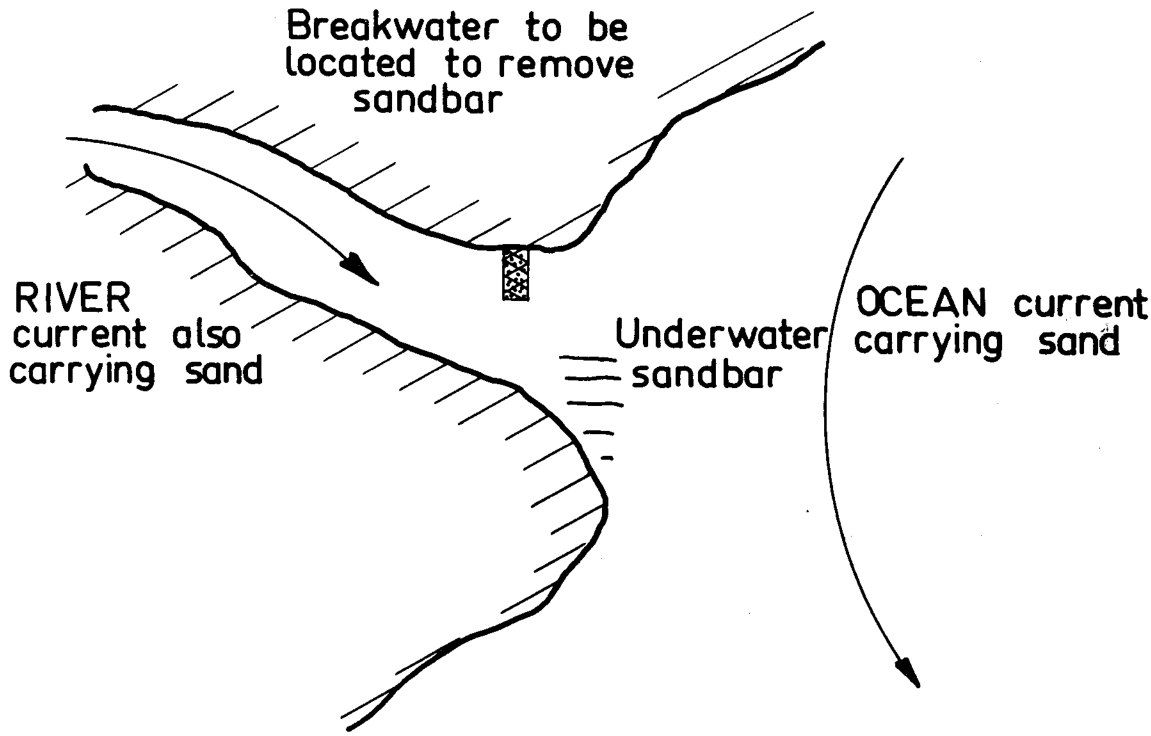


FIGURE 16.7
HARBOUR SILTING SYSTEM

16.9 The Economic System

Possibly the most complex system which mankind has devised is that which regulates the flow of goods and services in our community. The system is clearly one of great complexity, and of great importance, and one which has been the subject of intensive study especially in the present century.

The dynamics of the system depend very much upon human emotional reactions which are complicated and unpredictable, although when statistical averaging methods are employed it becomes possible to some degree to define in an approximate way the "average person".

Our knowledge of methods of determining the control decision sequences are very rudimentary indeed. Even to-day it is not certain how to prevent economic crises and the continuous oscillations of booms and slumps. It is of course appreciated that the Government, through monetary control, has an essential part to play in keeping the system functioning smoothly.

Because of the complexity of the system, it is not possible at present to establish very satisfactory abstract models, either in mathematical form or using computer simulation. Rather, the system itself is used for experimental purposes. Financial allocations and the tax rates are determined for yearly periods and the effect of these decisions upon the financial system is carefully observed. On the basis of the experience gained, the controls chosen for the following year are modified in the hope of achieving some more desirable performance in the year ahead.

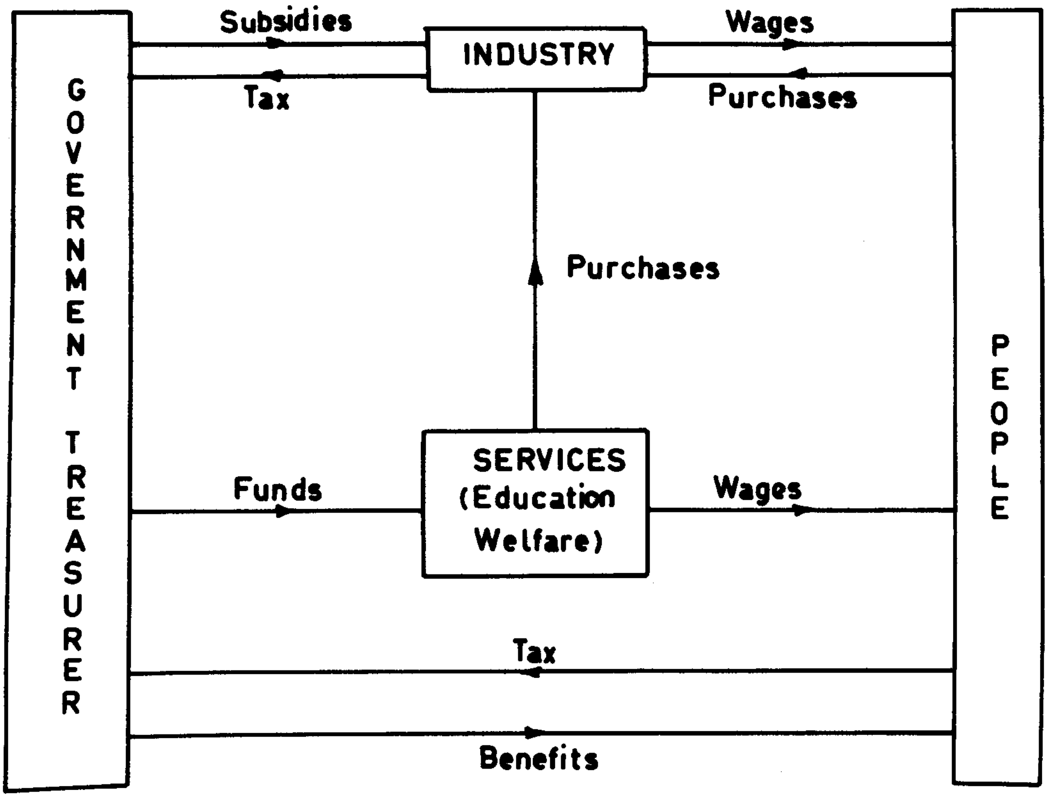


FIGURE 16.8
MONEY FLOW IN THE ECONOMIC SYSTEM

So we observe that in this very complex system, there is considerable lack of precision in our knowledge of the dynamics of these systems and hence of the effect which a particular control decision sequence might have upon the response of the system. Equally uncertain is the structure of the performance criterion which should be used.

The control system is one with very strong feedback at present. A simple change is made in the controls (fund allocations or tax rates) and the effects upon the system carefully observed, and used as a basis for deciding the next change in the control sequence.

	Satellite Altitude	Vehicle Pos.	Satellite Launch	Power	Traffic	Hydraulic	Economic
Models							
Precise mathematical model	√						
Approximate mathematical model		√	√	√			
Computer simulation					√		
Physical model						√	
System itself							√
Control decision sequence							
Precise without feedback	√	√					
Approximate with some feedback			√	√			
Rudimentary with considerable feedback					√		√
Single shot with no feedback						√	
Performance criteria							
Precisely defined	√	√	√				
Well-defined				√		√	
Poorly defined					√		√

FIGURE 16.9

SUMMARY TABLE

16.10 Exercises

Exercise 1:

Consider a simple system comprising a mass M moving on a horizontal surface in response to a control force u, and suppose that there is a viscous frictional force between the mass and the surface proportional to the velocity of the mass. Show that the state equations for such a system are of the form

$$x_1' = x_2$$

$$x_2' = -x_2 + u$$

where x_1 denotes the displacement

and x_2 denotes the velocity of the mass. Using an approximate numerical procedure calculate the response, namely,

$$x_1(t) \;, \quad 0 \quad t$$

and

$$x_2(t) \;. \quad 0 \quad t$$

to a control force

$$u(t) = \begin{cases} 0 & t \quad 0 \\ 1 & 0 \quad t \end{cases}$$

assuming that at time $t + 0$

$$x_1(0) = 0$$

and

$$x_2(0) = 0$$

In solving this problem use increments of time $t = 0.2$,

and note that initially

$$x'_1(0) = x_2(0) = 0$$

and

$$x'_2(0) = x_2(0) + u(0) = -0+1 = 1.$$

As would be expected results will show that a limiting velocity will be achieved, having a value of unity, at which velocity the friction force equals the driving force.

Exercise 2:

Regard a joinery factory as a dynamical system, and draw a block diagram indicating the state and control variables. Also, endeavour to define an error index or performance criterion in terms of the state and control variables. It is appreciated that this is in fact a very difficult question, but it does provide a useful vehicle for thought on the dynamics of large systems. A study of this problem will give much insight into the control of human created systems.

16.11 Suggested Further Reading

Further reading into the more quantitative aspects of this subject may be undertaken in the interesting book, *"Introduction to System Dynamics"*, J.L. Shearer, A.T. Murphy and H.H. Richardson. Addison-Wesley Publishing Company, 1967.

CHAPTER 17

Communication Systems

'Over the mountains
And over the waves ...'
Anon.

Chapter 17

COMMUNICATION SYSTEMS

17.1 Types of Systems

Communication systems can be conveniently classified into two broad classes.

Class 1. Transportation of information. Here the principal objective is to convey information between two points distant in space (Fig. 17.1)

Class 2. Extraction of information. Here the objective is to extract the information about the environment and to act on it (e.g. radar defence system, space projects etc.)

The principal objective of a communication system of Class 1 is to convey information from one point of space to another using various transmission media and devices. A telephone network is a typical example belonging to this class. A radio-telephone is another one, although here the medium of communication is different. With some systems, various ancillary apparatus is needed (Fig. 17.1) to perform a variety of functions such as encoding of the input data into a form suitable for transmission along the channel. At the receiving end the signal is processed by means of a suitable decoder into a form required at the destination.

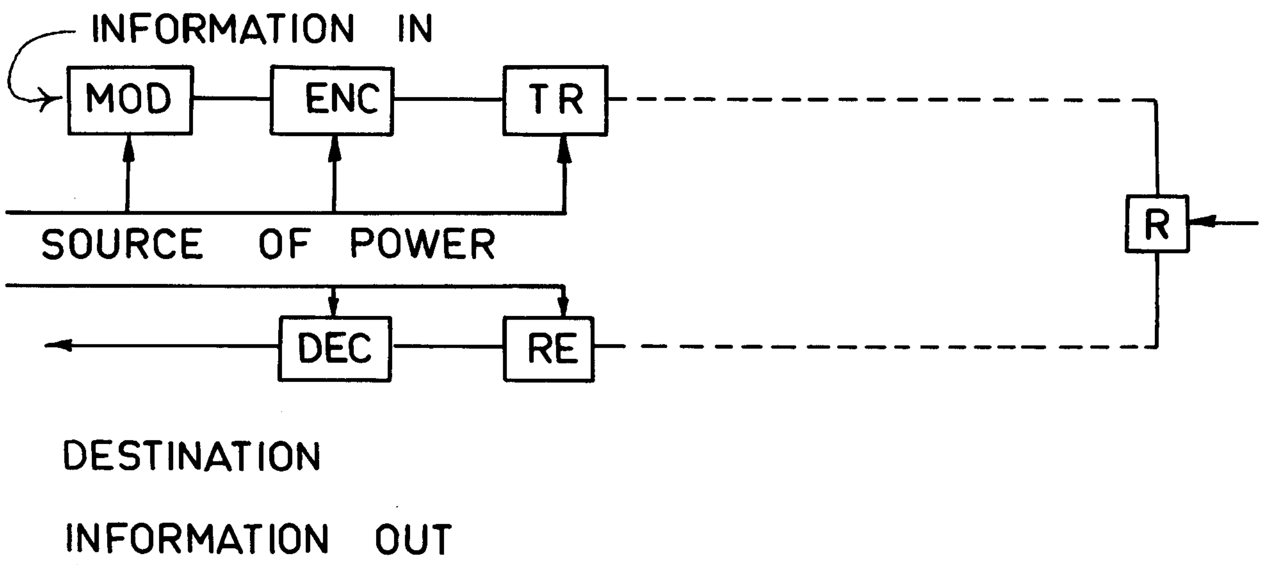


FIG. 17·1. TYPE 1 COMMUNICATION SYSTEM

Radar is a communication system in a somewhat different sense. The source is a high power directive transmitter emitting signals of known form. The receiver, usually placed alongside the transmitter, detects the signals reflected by distant objects, such as an aircraft, and the signal is invariably accompanied by noise, as well as being distorted in form due to numerous contributing factors. With such a system the information received is really not the information transmitted, since the transmitted signals (the transmitted data) is a periodic series of pulses and, as such, contains no information. The receiver is required to trace a different kind of information, information on the position of the target in terms of distance, bearing and elevation. This information is obtained in an indirect manner by correlating the received pattern of pulses with that which has been transmitted. Clearly, in view of the previous discussion such information can be extracted with precision from a noisy signal by various artifices such as integration or correlation.

In effect a radar system consists of two sub-systems (transmitter and receiver) interconnected by a data link (radio path) in such a way as to determine the information about the position of the target. It is an example of a relatively simple system.

One objective of radar measurements is to determine the position of an enemy target so that subsequently through human actions defensive means, such as anti-aircraft guns, can be brought into operation to intercept the enemy target. The complete operation consists of a radar system linked to a defense system using human intervention, with all the attendant disadvantages. The next step in improving the overall performance of such a defense system would be to link the radar system directly to anti-aircraft guns by means of a date link, thus permitting a direct control of the anti-aircraft guns in accordance with the data received by the radar system under an overall control of a suitable computer. Fig. 17.2 shows the schematic of the system, provision being made for direct human control if need be.

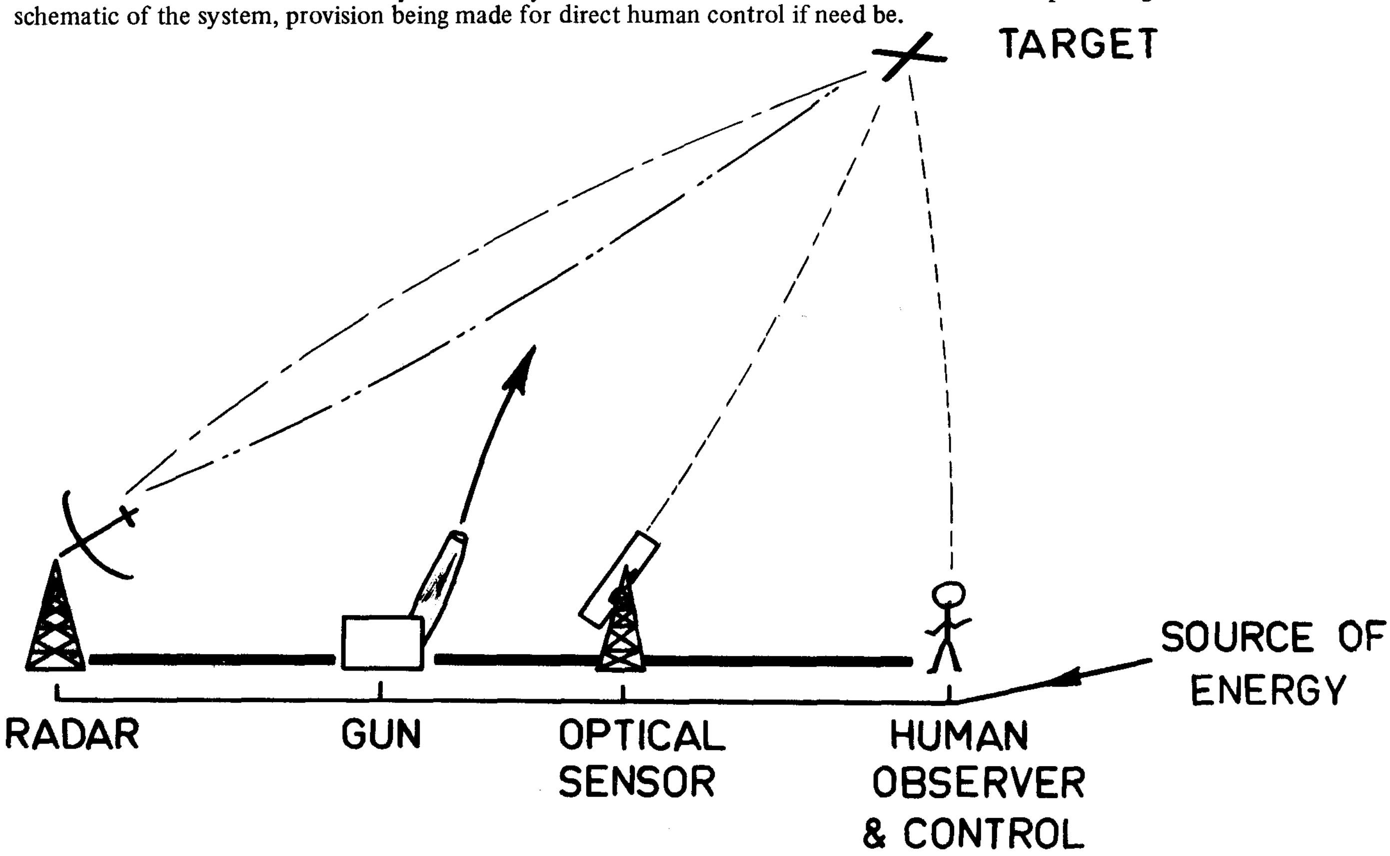


FIG. 17·2. TYPE 2 COMMUNICATION SYSTEM

Another, though less obvious, example of a communication system is that of a scientist together with his instruments engaged in an investigation. Here, clearly the scientist is transmitting various signals into the physical environment and then he is observing the response of the system, and by interpreting the response of the system in relation to the inputs, he derives information about the physical system. Such a system can be looked upon as an embodiment of a communication system of class 2 with the data of measurements forming the output. It is at this juncture that we are likely to have a conflict between the two notions of information: the mathematical and the semantic. As far as communication theory is concerned we can analyse the system in minute detail. We can also determine the limits of significance of the data (be it in digital or analogue form), we can determine the maximum permissible flow of data and even suggest means for improving the system so as to increase the accuracy of the data. But we cannot say anything about the value, meaning or significance of the data as far as the scientific observer is concerned, for this would be semantic information. The data could be a meaningless thermal vibration waveform or the key signature of the DNA chain, this however, would be of no interest to the communication theory, it is outside its province.

To put this important aspect in yet a different way. The output data of an elaborate scientific equipment will wander aimlessly even with the help of the communication theory and the most elaborate computer, but when processed by skilful thoughts of a scientist, it acquires meaning and beauty in the form of the patterns created.

17.2 The Characteristics of Systems.

Communication systems can exist in a variety of forms, but intrinsic to them all are three fundamental characteristics. These are

1. Usually the source of information and the destination are physically separated. Therefore an instantaneous transmission of the information is impossible on account of the finite velocity with which signals travel (the highest velocity being the velocity of light). This imposes fundamental limits due to delays in transmission.

2. Because of imperfections in the system signals will be mutilated by the channel as well as by the ancillary apparatus. There are in fact definite limits on what can be achieved.

3. All apparatus and processes of communication are subjected to external interferences and noise (see Chapter 2), on account of which a truly unequivocal communication is an ideal which cannot be realised in practice. Errors in communication are therefore inevitable.

The aim of the communication engineer is to design a system so as to transmit a given class of information bearing patterns with minimum equivocation while optimising a required number of key parameters, such as cost. Naturally, the design is a compromise and communication theory can help to strike a better bargain.

17.3 Communication systems: A brief outline of developments.

When discussing the history of electrical communication systems one needs to be careful in defining what electrical communication means. We know from historical accounts that electrical communication systems, as we know them today, were preceded by various other communication systems. Thus for example, the electrical telegraph system was anticipated by such systems as the semaphore system. Indeed, the two systems have a great deal in common, in that both use a method of encoding the individual letters of the message to be transmitted into a system of symbols for transmission from one point of space to another. The only difference seems to be that in one case one would use radio waves as the carriers of information whereas in the other case the symbols would be generated by mechanical means and observed by optical means from a distant point. In essence however, in both cases the transmission of symbols is taking place by means of electromagnetic waves (with radio systems it is the radio waves which carry the information, whereas with the semaphore system it is electromagnetic waves in the visible part of the spectrum (See Fig. 17.3)). While discussing the merits of communication systems against the background of history it may be instructive to observe that the early heliograph system had a great deal in common with simple communication systems of today using laser sources.

The principal disadvantage of early communication systems of the type of semaphore or heliograph, was their inability to communicate at other than very slow rates, as well as their unreliability. The systems were capable of being used only under good weather conditions and would fail totally when the visibility due to fog or heavy rainfall was restricted.

Electrical communication systems can be said to have started with the invention of the electrical telegraph system which seems to have taken place simultaneously on both sides of the Atlantic: in U.S.A. by Morse and in England by Wheatstone. The essence of the idea here, was to transmit electrical signals in the form of pulses of defined duration, and by encoding individual letters of the English alphabet into a sequence of dots and dashes it was possible to send messages along a pair of wires (or even one wire and the earth). As such, the system was an early example of encoding from one finite set of symbols into another.

Telephony is an example of an anologue communication system. The system came into being through the invention of two devices: a transducer for transforming the soundwave into electrical impulses i.e. a microphone and a transducer for converting electrical signals into soundwaves i.e. an earphone.

Communication systems of today relate to one or the other of the above described communication systems. The differences such as there are relate to sophistication in equipment and different methods of transmitting the signals. In essence however, there are only two methods of transmitting signals: 1. By radio waves, i.e. by free space communication, 2. By various types of waveguides such as transmission lines and coaxial cables.

Fig. 17.4 summarises some of the principal characteristics of different methods of conveying electrical signals in use today.

Taking the world as a whole, the most frequently used communication media for telephone traffic (when expressed in terms of circuit miles) are the open-wire and balanced-pair cables, which incidentally are the oldest means of communication. Originally, however, one pair of wires was needed for every telephone channel and consequently the cost of the system was essentially proportional to its capacity. By providing a number of separate carrier frequencies on each pair of wires and modulating the voice of separate subscribers onto the different carriers, a number of simultaneous telephone conversations can be transmitted along each pair of wires. At the receiving end, the different telephone channels are sorted out by the use of filter which separate the different carrier frequencies and with them the intelligence transmitted. This is known as multiplexing.

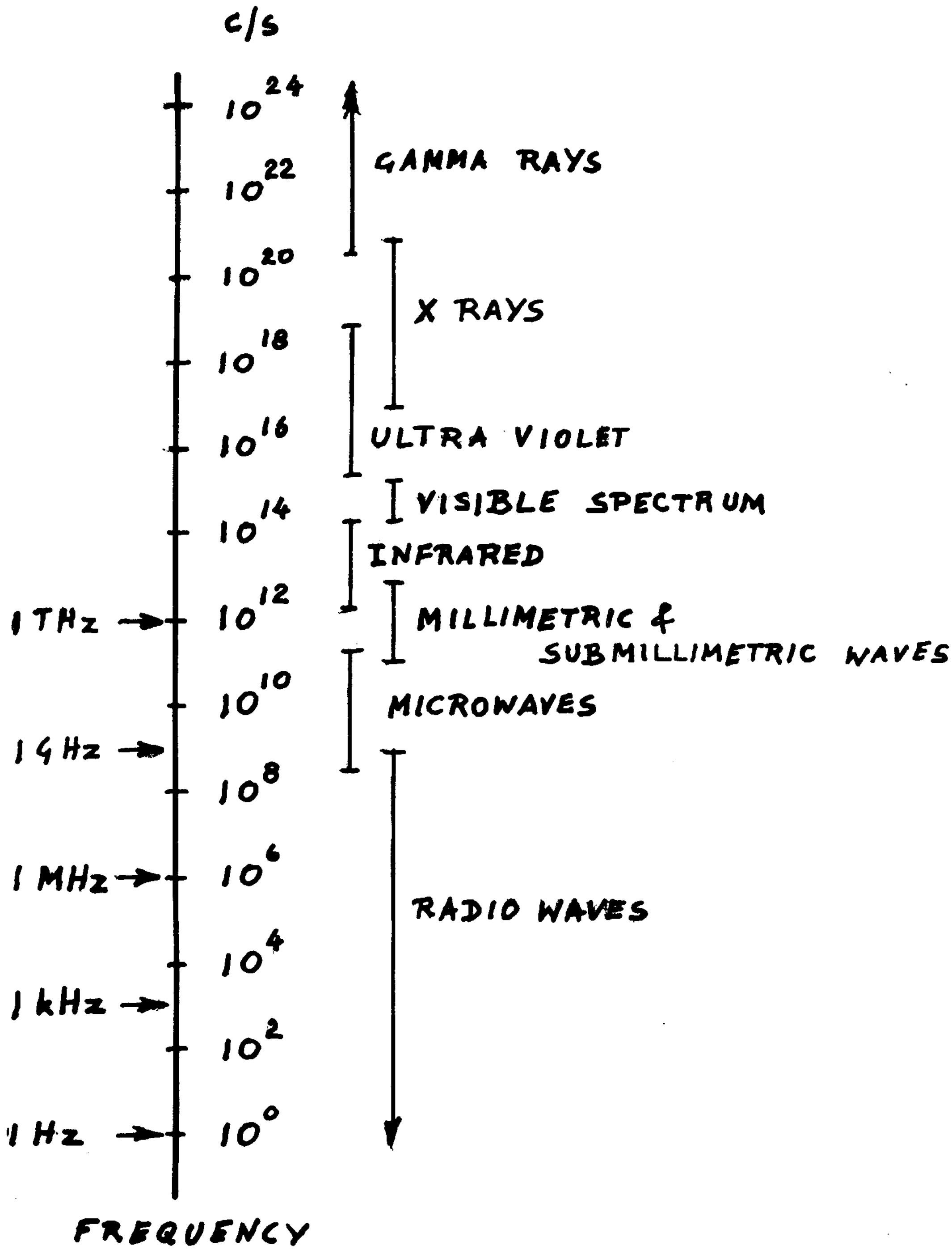


FIG. 17.3. ELECTROMAGNETIC SPECTRUM.

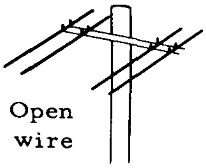
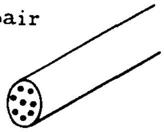
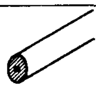
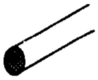
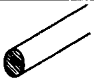
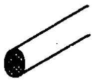
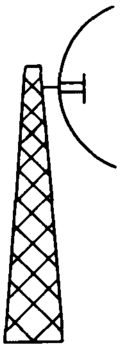
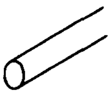
Type		Operating frequency	Channel capacity	Repeater spacing / attenuation
 Open wire	One pair	36 to 84 kc/s (go) 92 to 140 kc/s (return)	12 + 3 + (1) = 15 + (1)	60 to 70 miles (0.4 db/mile)
	Maximum 16 pairs	As above	240 + (16)	
Balanced pair 	One pair	12 to 252 kc/s	60	12 to 30 miles (4 db/mile)
	Maximum 24 pairs	As above	1440	
Coaxial G. B.		Up to 4 Mc/s	960 (One super channel)	6 miles
		Up to 12 Mc/s	3 super channels	3 miles
Coaxial U. S.		Up to about 3 Mc/s	720	8 miles
		L3 system	1 TV channel + 600 speech channels	4 miles
Single wire transmission line		About 100 to 1,000 Mc/s	Probably a few TV channels	About 2 to 10 miles
Microwave links 		Below 500 Mc/s	Maximum of 60	40 to 50 miles
		500 to 1,000 Mc/s	Up to 120	40 to 50 miles
		2,000 Mc/s	240×6 or $1\,\mathrm{TV} \times 6$	30 to 40 miles
		4,000 Mc/s	600×6 or $1\,\mathrm{TV} \times 6$	25 to 30 miles
		6,000 to 8,000 Mc/s	600×6 or $1\,\mathrm{TV} \times 6$ Maximum $2\,\mathrm{TV} \times 6$	25 to 30 miles
		11,000 Mc/s	Less than 600	Less than 20 miles
Long haul waveguide (H_{01})		30,000 Mc/s up to about 100,000 Mc/s	1,000 super channels ≡ several hundreds of TV channels ≡ several hundred thousands of speech channels	20 to 40 miles (attenuation 2 to 4 db/mile)

Fig. 17.4 Communication Systems and their capacity

Multiplexing cannot, for a number of technical reasons, be extended indefinitely by cramming more and more telephone conversations along one pair of wires. The chief reason is that for an acceptable reproduction of human speech, about 4KHz bandwidth is needed. Transmission of, for example, 60 simultaneous telephone conversations, necessitates 60 x 4 = 240 KHz and, the higher the frequencies involved, the larger the attenuation, cross-talk between adjacent pairs, and signal distortion. To keep these effects within bounds, it would be necessary to reduce the repeater spacing; consequently the total number of repeaters needed would have to be increased, thereby augmenting the cost of the complete installation.

The design of a communication system is a careful balance between numerous, often conflicting, factors to minimize the total cost for a given performance. With modern equipment, up to about 15 simultaneous telephone conversations can be transmitted along an open pair of wires and 60 along each pair of wires of a balanced-pair cable system. The permissible number of pairs of wires in any one system is also limited by various factors, principally the maximum permissible cross-talk between the pairs. Thus, for example, the present practice with balanced-pair cables is to provide 24 pairs each having a capacity of 60 telephone channels giving a total capacity of 1,440 telephone channels and a repeater* spacing of the order of 20 miles.

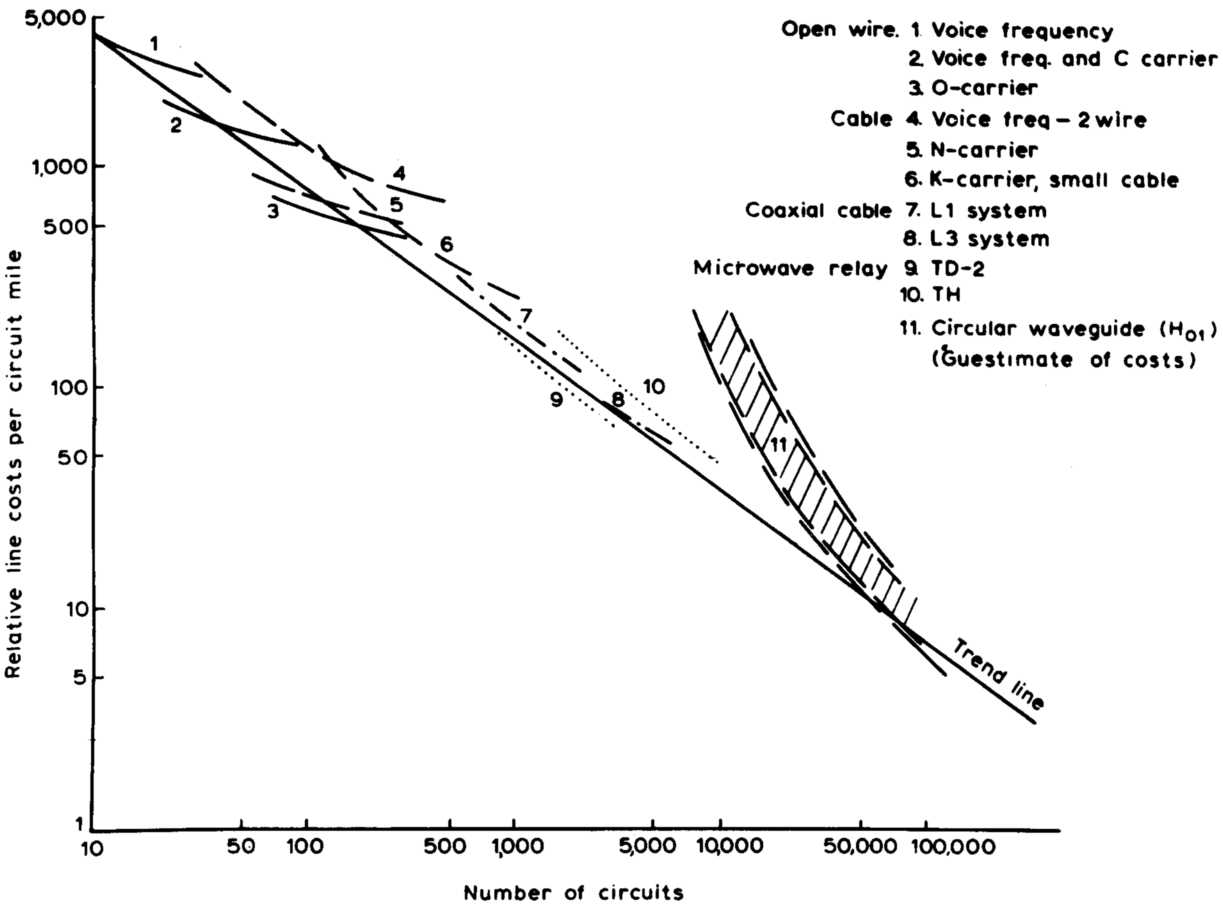


Fig. 17.5 Trend Line of Line Costs

This is just one illustration of the aims of engineering design: given the objective, to use existing skills and techniques so as to minimize the cost of achieving the objectives. In the above example of the balanced pair, the 1,440 telephone channels could, in principle, be provided by simple voice channels on 1,440 pairs of wires. Instead, using multiplexing techniques the same capacity can be provided on only 24 pairs of wires. Clearly, in effect, the cost of line is traded for the cost of somewhat more complex terminal equipment and repeaters – and on balance the transaction is well worth while.

* A repeater is a device inserted along a transmission line to amplify and correct the signals carried.

As a general rule, the heavier the communication traffic on a given route the greater the likelihood that a system of great technological complexity will prove to be economically successful. This is clearly supported by long years of experience. Fig. 17.5 shows the general trend of cost of communication (per channel) as a function of density of traffic for different types of communication systems.

There are many other factors entering into the design of communication systems (such as accessibility to repeater stations, climatic conditions, established practices, human prejudice, etc.) which need not be discussed here, but one additional factor must be considered: requirements for other than telephone transmission.

Since the Second World War there has been a great expansion in television. Whereas before 1939 there were only a few countries where television transmission had been attempted, today in many countries several independent and simultaneous television programmes are in operation and are served by a wide distribution network (e.g. Euro-vision). Each television channel requires for transmission as much bandwidth as 1,000 telephone channels and this bandwidth must be provided in a single communication medium. Clearly, open-wire or balanced-pair cables are unsuitable, because a single pair of wires cannot transmit efficiently such a wide bandwidth.

However, a modern coaxial cable or a microwave link can easily accommodate several television channels in addition to the telephone traffic. These are the wide-band communication systems of today. For wide-band applications it is usually microwave links that are economically attractive. But, unlike coaxial cables, these use for the carrier frequency, microwaves in the frequency range thousands of megahertz. (See Fig. 17.3). Furthermore, a microwave link is an example of an 'open' communication system in that, like radio, it utilizes free space as the vehicle for information transmission, and to avoid mutual interference careful planning is necessary as regards the frequency allocation. This frequency allocation is governed by international agreements, and in many highly populated areas free space is at a premium. In the next decade or two it is very likely that we shall witness a saturation of free space, particularly as this is more urgently needed for communication with moving objects, such as ships and aeroplanes.

Coaxial cables are at present the backbone of wide-band large capacity communication systems. With growing demands on communication channels we must look forward to new means of communication to meet the demands of the future.

17.4 Some future possibilities.

In the last few decades we have witnessed an unprecedented rise in the demands on communication channels. There is every indication that the increase in demands of communication channels will continue at a rapid rate. These demands are so real that a phrase "explosion of communications" has been coined to describe the situation. Nowadays many thousands of speech channels are required to provide for communication between large cities. In addition television requirements have risen enormously, and these demand channel capacities many times in excess of those needed for speech communication (one television channel $\approx$ 1,000 telephone channels).

We can already foresee new demands for communication channels for various reasons which are apparent. There are increased needs for facsimile transmission and data transmission for various business purposes such as in commerce or in banking, in addition to the growing requirements for communication with computers. For every day use video-telephone, i.e. a telephone which combines transmission of sound waveforms as well as pictures has already been developed and could be put into operation provided that the greatly increased demands on communication channels could be satisfied. We foresee a great increase in the use of picture as well as speech transmission for purposes such as conferences between managers of different branches of the same company or for direct communication of technical information within the structure of a company, and also for transmission of technical and other information from central computer memory store to various customers.

Indeed, in the years to come, there should be no need for people to spend hours in patent or technical libraries searching for information needed in their work. Such functions can be performed much more economically by direct communication with a computer. The computer in response to a request from a customer would automatically search the stored information and display on a television screen the information requested by the customer. Indeed, there seems to be no end to the varieties of services which could be provided in the years to come by direct communication with computers. At the moment however, such possibilities must wait their implementation until the problem of providing large numbers of communication channels is solved.

When assessing the future possibilities for new communication channels, it is useful to examine the bounds of possibilities arising from fundamentals of physics. First of all, it is necessary to accept that in the future it will not be possible to make more extensive use of radio-waves for point to point communication, in that radio-waves will be needed to a greater extent than ever for communication with moving objects such as aeroplanes and ships.

There are however, distinct possibilities in the use of satellite communications using UHF frequencies, in that a properly planned satellite communication system would offer a more economical use of the radio frequency spectrum. There are however, definite limits to what can be achieved, and it would appear, on the evidence before us, that satellite communication systems cannot offer a long term solution to the problems ahead of us. The possibilities such as exist would seem to lie in greater exploitation of higher frequencies than UHF. Here we have in mind (see Fig. 17.3) the utilisation of the frequencies

well beyond the microwave part of the spectrum, such as the millimetric spectrum (frequencies corresponding to the band (30 to 300 GHz), or the sub-milimetric part of the spectrum (300 to 3000 GHz). It is also possible that frequencies even higher than those could be used for communication. In particular with the advent of lasers the exploitation of the infrared frequencies and the frequencies corresponding to the visible part of the spectrum (the whole of these two spectra is known under the name "optical spectrum") has now become a possibility.

We can now generate optical frequencies with an efficiency approaching that with which we can produce radiowaves. Thus, a point to point optical communication is one possibility, but there are even more exciting possibilities lying ahead of us in systems utilizing guided optical waves in suitable waveguides, not unlike radiowaves which can be guided by coaxial cables. (See Reference 5).

Millimetric as well as sub–millimetric waves offer exciting possibilities. Here we have in mind guiding such waves inside metallic tubes of suitable construction (See Reference 4). With such systems, it is estimated that it should be possible to transmit several hundreds of video channels in addition to many thousands of speech channels.

The recent advances which have been made in materials science and the technology of microminiature circuits have opened up new possibilities in all-digital integrated microwave communication systems. There is no doubt at all that in the next two decades, we shall witness great progress in such communcation systems.

Looking further ahead, there seems to be no reason why communication on a wider scale should not be made possible in the years to come. To provide hearing for the deaf and vision for the blind are distinct possibilities in the present day technology. But while these possibilities are under investigation, new possibilities emerge in communication by extending the capacity of the remaining human senses. Here we have in mind remote sensing by touch, smell or even taste. The basic issues here are the invention of suitable transducers, in the same way as telephone communication was made possible through the invention of the microphone and the earphone.

On another forefront, work is currently proceding to try to expand communication theory into the domain of semantics. This would seem to be a field full of difficulties ahead but nonetheless, once we learn to understand how to communicate semantics we shall have opened new possibllities in communication. Some philosophers even go so far as to foresee possibilities of direct communication of meaning as well as human feelings, and even the basic ideas behind human thought. While it is true to say that such possibilities are not, at the moment within the possibilities of present day technology, they nonetheless cannot be excluded from possible exploitation in the future. But, before such systems become a reality, we should do well to examine the likely effect which they might have on human societies.

17.5 Suggested reading material

1. Singh, J. "Great Ideas in Information Theory, Language and Cybernetics"
(Dover, 1966)

2. Karbowiak, A.E. "Theory of Communication"
(Oliver and Boyd, 1969)

3. Beck, A.H.W. "Words and Waves"
(World Univ. Lib. 1967)

4. Karbowiak, A.E. "Trunk Waveguide Communication
(Chapman & Hall, 1965)

5. Karbowiak, A.E. "Optical Waveguides"
(In Advances in Microwaves Vol. 1, Academic Press, 1966).

Chapter 18

ELECTRIC POWER AND COMPUTERS

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The linking of "power" and "computers" might seem a little unusual since it is only within the last 15 years that the theory of discrete control systems has been actively developed. Only comparatively recently have digital computers been used for control purposes and it would be correct to say that their application to power systems has barely commenced when viewed in the light of their true potential.

"Systems Engineering" and the "Control of large Systems" are topical subjects destined to play important roles in the electric power industry.

History has shown that large systems evolve or grow from small beginnings. Their control by man develops through experience. At various stages of the growth of the system crises arise and problems which emerge must be solved before the system can be safely enlarged.

It is therefore relevant to make some mention of the development of power systems and the problems which have arisen.

18.1 Early History

The first alternating current system in America using transformers was installed at Great Barrington in Massachusetts in 1886 by William Stanley. The Siemens alternator rated at 500 volts, 12 amperes was imported from England by George Westinghouse. He had secured all rights in the United States for the Gaulard and Gibbs transformer. Stanley designed and constructed a number of transformers wound for a 500 volt primary and a 100 volt secondary, six of which were installed in the town of Great Barrington. It is interesting to note that the distance from the generator to the centre of the town was 4000 feet. Since the electric power industry of today is predominantly alternating current we could regard the Great Barrington installation as the beginning of the modern electric power system or "integrated power pool".

18.2 Sources of Energy

The sources of energy for the generation of electric power may be classified as follows: –

1. Solar Energy:
   - (a) Directly by photocells or steam raising for the generation of electricity.
   - (b) Photo-synthesis; to produce combustible vegetable matter* and fossil fuels such as coal, petroleum and natural gas.
   - (c) Wind power
   - (d) Precipitation hydro-electric power
2. Tidal hydraulic power
3. Geothermal power
4. Chemical energy (not involving combustion)
5. Nuclear energy – fission or fusion.

Schemes have been developed to focus the rays of the sun by large mirrors on a boiler for steam raising. However, it has been estimated that solar radiation falling on the whole of the land surface could satisy all energy requirements if conversion could be effected at an average efficiency of 2.5 percent, but unfortunately the technology for its achievement does not exist today. Even if it did exist, the cost wohld be prohibitive.

Attempts were made at Abadan on the East Coast of Africa to utilize the difference in the temperature of the sea water at the surface and at a depth of a few hundred feet. Wind generators have been installed in various parts of the world but it has been concluded that our present technology does not enable us to convert the large amounts of energy in the movement of the earth's atmosphere.

* The Townsville Regional Electricity Board accepts generation of electric power from sugar mills using begasse as fuel.

Geothermal power remains largely an unknown quantity. However Geothermal stations exist in various parts of the world; for example, at Torpo in the North Island of New Zealand, approximately 200 MW. are generated from a geothermal source.

Chemical energy in the form of fuel cells may yet assume great importance for small sources of power but for economic and other reasons it will not play a major role in large scale power generation.

Hydro-electric power is recognised as an important source in many countries, including Sweden, United States of America, Canada, Britain, Australia, etc. In Tasmania, for example, the entire generating plant is hydro–electric However, in the long range future, it has been recognised that precipitation hydraulic power can provide only a small proportion of the total energy requirement.

Tidal power has been recognised as an importantsource of energy for many years. Coastal dwellers have been trying for centuries past to harness the energy resulting from the ebb and flow of the tide. Paddle wheels set in motion by the sea water have served to work grain crushers. Such low capacity tidal mills were to be seen on the Breton Coasts of France as early as the 12th century. One or two were in operation until recently in the Rance River estuary near St. Malo and elsewhere in Brittany. The tide rises and falls twice in every 24 hours and 50 minutes which is the apparent period of rotation of the moon. It reaches a maximum every two weeks when the moon and the sun act in conjunction; that is, when these two bodies are pulling together we have what is termed a "spring" tide.

Unfortunately there is a relatively limited number of locations throughout the world where the tide amplitude is large enough to justify a tidal power station. For example, in the Bay of Funday on the Atlantic Coast of Canada the tide amplitude exceeds 50 feet and a series of tidal plants has been proposed at the Passamaquoddy site.

Alaska has sites with tidal amplitudes exceeding 33 feet. In the Gulf of California near the mouth of the Colorado River the tidal variations reach 30 feet. The Severn estuary has a tidal amplitude in excess of 45 feet and Argentina, India, Korea and parts of the northern coast line of Australia have good sites as does the Russian North Coast.

In France there are several good locations with a tidal range of 40 feet or more, e.g. Brittany. The planetary system sets the rhythm of the seas, but only the sea, with the constraints of the geography of land masses, decides the nuances and values of this cadence. The simplest tidal power plant involves damming an estuary or other suitable location, admitting tidal water to the basin thus formed, closing the sluice gates until the tide has eased and then allowing the basin water to flow back to the sea through turbines to generate power. More sophisticated schemes involve turbining during filling and pumping in off-tide hours to raise or lower the level of the main or adjacent basins. Combinations of these schemes permit energy to be extracted at times more suited to the industrial demand for power. The turbining and pumping functions have been combined into a single piece of machinery which can function as a turbine with water flowing in either direction. In other words its unique design enables it to perform the same tasks as four separate machines.

The first tidal power plant is the Rance River project which was constructed by Electricité de France. It is located about two miles upstream of the town of St. Malo. The tidal variations are of the order of 40 feet and 24-10 MW units are being installed, the first of which went into operation in 1966. The average head of water is about 25 feet.

At the Passamaquoddy site in the mouth of the Bay of Funday it is planned to build two 500 MW plants.

The changing pattern of energy sources is highlighted by the fact that in 1947, coal supplied roughly half of all energy consumed; by 1965 it had dropped to about 1/5th of the total and this proportion continues to fall.

Although the consumption of petroleum, natural gas and coal will continue to increase absolutely, these fossils fuels will in the future account for a relatively smaller proportion of the total energy produced. The hope for the future rests largely on nuclear energy.

18.3 Generator Unit Size

The size of generator units has continued to increase rather rapidly over the past 20 years and naturally the question arises as to whether there is an optimum size of machine for a given power system.

Investigation has shown that the most economical method of power system expansion is to add generation units of 7 to 10 percent of the size of the system. This assumes that the investment cost per kw of generating units continues to decrease with size and the forced outage rate for large generators remains at the present level. There is no indication that the maximum economic size of units has been reached and machines of 1000 to 1200 MW are now in service.

18.4 Transmission of Power

The main factor determining the adoption of alternating current has been the ease with which it can be converted from one voltage to another by means of a "transformer". At the time when the early choice between a.c. and d.c. was being made,

the induction motor as invented by Tesla, provided the prospect of an extremely simple industrial machine. In spite of the decision in favour of alternating current (a.c.) many cities operated on direct current supplied by rotary converters and mercury arc rectifiers; the changeover of the Australian capital cities to alternating current came as late as the 1930 - 1945 period. Variable speed could be provided more easily by direct current machines for lift drives and certain industrial applications. A key role in the generation, transmission and utilization of energy is played by the power transformer. As with generator unit sizes, transformer capacities have increased and should reach 1100 MVA by 1970 Without the transmission of power at ever increasing voltage there could not have been the vast increase in consumption of power for industrial expansion.

Historically, the first long high voltage transmission line was built in 1914 by the Southern California Edison Company. The 243 mile line from the Sierra Nevada Mountains to Los Angeles operated initially at 150 kV and in 1924 the voltage was raised to 220 kV. In 1936 the Los Angeles Department of Water and Power constructed the 265 mile 287 kV Boulder lines feeding Los Angeles. Sweden's 593 mile 380/420 kV line was completed in 1952. More recently, the Hydro-Quebec of Canada has built a 700/735 kV 400 mile system to develop the vast hydro-electric potential of Northern Canada and it is believed that a.c. voltages will reach 1000 kV in the not too distant future.

Naturally the energy sources near the load centres were the first to be developed. However, it soon became necessary to build power stations at considerable distances from the industrial towns and problems arose over the stability of the electrical system. Higher voltages were adopted to increase the capacity of the transmission lines.

Economical extra-long distance point to point transmission lines are becoming difficult to design because of the high cost of the compensation or resonance equipment required to maintain stability. Two systems, inter-connected through an a.c. line must run in synchronism, i.e. the rotors of their generators must remain fixed within certain limits with respect to each other. Sufficient stability is provided only if this angular displacement or the phase shift does not exceed 30 electrical degrees. On a 60 c/s system, a 250 mile line with surge impedance loading has a phase angle of approximately 30°. To go beyond this length requires complete compensation for miles in excess of 250 with consequent cost penalties.

However, communications engineers encounter long lines and are aware of the interesting phenomenon which occurs for electrical lengths between ½ and ¾ wave length (or 180° to 270°). The system is as stable as one operating in the first quadrant, namely 0° to 90°. The most attractive feature of the half-wave length system is that the cost per unit length decreases as the line length increases. The transition point between the conventional and the half wave length system is 900 miles.

TABLE 18.1

SUMMARY OF HIGH VOLTAGE TRANSMISSION SYSTEMS

Location	Installation Date	Rating kV	Rating MW	Type of transmission and distance	Remarks
Gotland (Sweden)	1954	100	20	60 miles-cable	Swedish mainland to Gotland Island
English Channel (France-England)	1961	±100	160	40 miles-cable	Interconnecting English and French systems
Donbass-Volgograd (U.S.S.R.)	1964	±400	750	300 miles-overhead line	
New Zealand (Cook Strait)	1965	±250	600	25 miles-cable 360 miles-overhead line	Transmitting power from the South to the North Island
Japan (50-60 c/s)	1965	2 x 125	300		
Konti-Skan	1965	250	250	54 miles-cable 59 miles-overhead line	Connecting Denmark and Sweden
Sardinia - Italy via Corsica	1966	200	200	61 miles-cable 217 miles-overhead line.	

Alternatively, much attention has been paid to the use of high voltage direct current (d.c.). In principle such lines have no stability limitations. The cost of the basic d.c. transmission line is always less than that of the a.c. line. For submarine installations the difference is particularly pronounced as the sea may be used as an earth return. For a given capacity, an overhead d.c. transmission line costs about two thirds of the a.c. line. Usually it is not a simple case of a d.c. source with d.c. transmission to a d.c. load; in general the d.c. line connects two a.c. systems.

A summary of high voltage d.c. transmission systems is given in Table 18.1

It is worth noting that the ground is an excellent return conductor for a d.c. line. With a.c., the inductive effect forces the ground current to follow the line route near the surface. By contrast, the d.c. ground current follows the path of least resistance, penetrating the good conducting interior of the earth. The only notable ground effects are in the vicinity of the earthing electrodes.

High voltage d.c. transmission would have been adopted much earlier had it been possible to construct suitable high voltage d.c. generators and/or motors. Important research with this end in view was carried out by Noel Felici at Grenoble, France. He concluded that the Whimshurst machine was not efficiently designed and proceeded to develop a new form of electrostatic machine. He foreshadowed that the Felici generator might ultimately produce enough power at a voltage suitable for d.c. power transmission. At the receiving end the line could be coupled to the a.c. network through an electrostatic motor driving a conventional synchronous alternator. However, it appears unlikely that the conversion equipment as developed by A.S.E.A. of Sweden (see figures 18.1 and 18.2) will be replaced by rotating machines.

During the last few years efforts have been directed towards the development of solid state high voltage d.c. terminal equipment in the form of thyristors. The feasibility of connecting thyristors in series has been conclusively demonstrated. The main tasks confronting the groups studying thyristor applications were the problems involved in gating series strings of thyristors as well as the problems of equal voltage distribution during switching. It is now possible to utilize the advances in solid state control and logic in order to take full advantage of the fast response times of thyristors as compared with the mercury arc devices which convert from a.c. to d.c. or invert from d.c. to a.c. With these developments the future of E.H.V. d.c. interties appears to be assured.

As a result of research and satisfactory experience in operation, high voltage d.c. lines are becoming recognised as the method of intertying large power systems. The Germans have begun work on a 400 kV system to transmit 60 MW 75 miles from Berlin to a terminal on the Elbe. Design work is in progress for a ± 750 kV 5250 MW d.c. transmission intertie from Kazakhstan to Tambov a distance of 2500 km and it is expected to be in operation in 1970.

The giant d.c. intertie of two 800 kV lines rated as 2700 MW from the Columbia River to the Hoover Dam and Los Angeles will have the highest voltage, largest capacity and be the longest in the world.

Electricity consumption has been doubling every eight to ten years and in most countries a doubling of the generation capacity is being planned for the next ten years. Although an annual growth rate of 7% is normal, the U.S.S.R. electrical energy production was 11% greater in 1965 than 1964, and the vast land masses of European and Asiatic Russia are being interconnected by long distance extra high voltage interties.

The tremendous developments of the Pacific North West Coast are typical of the pressures on the power industry brought about by the doubling every ten years. The Bonneville Power Administration serves an area of 290,000 square miles with 21 existing dam sites. The main source of its hydro-electric power is the Colorado River. The harnessing of the River did not commence until 1933 with the Rock Island Dam near Wenatchee, Washington. The Bonneville Dam was completed in 1939 and the Grand Coulee in 1941. Since then numerous dams have been constructed. The potential capacity of the Columbia River is far from exhausted and under an agreement between the U.S.A. and Canada three dams will be constructed in Canada namely, Duncan being completed in 1968, Arrow Lakes to be in operation in 1969 and Mica scheduled for 1973. The U.S.A. will construct Libby Dam in Montana. It is interesting to note that this dam will back water 42 miles into Canada. Apart from a third power station at Grand Coulee and a second at Bonneville dam, plans are well advanced for the Lower Granite, Asotin, Wells, Boundary, and High Mountain Sheep Dams. This vast electric power system is to have an intertie consisting of four major transmission lines from the major dams on the Columbia River to Los Angeles and the Hoover Dam. Two of the lines will each be 500 kV, 1000 MW capacity and the remaining two 750 kV, 1350 MW capacity each. The a.c. lines will run from John Day dam to Los Angeles. One d.c. line will connect to Los Angeles and the other to Hoover Dam. Other transmission lines will be added between Hoover Dam and Los Angeles.

18.5 Power Failures

At this point we need to pause and contemplate the real import of the rapid rate of expansion of power systems, namely the doubling in less than ten years. Consider the vast systems which have been constructed to the present date and try to assess the magnitude of the task involved in doubling the generating capacity together with the consequential network expansion accompanying it. At least we are aided by the improvements in technology, and the increase in the size of the units which will be used to construct the giant systems of the future. But as was stated earlier, systems evolve, generally from small

beginnings. With the expansion come problems; problems such as the North East power failure of November 9-10, 1965 which involved the massive outage of the states of New York, Vermont, Massachusetts, Connecticut and New Hampshire in the U.S.A. and Ontario in Canada. The period of the outage ranged from up to 2 hours in New Hampshire to a maximum of 18½ hours in parts of New York. So seriously did President Johnson regard the occurrence that he immediately appointed a Federal Power Commission. The urgency of the situation was underlined by the fact that in less than one month the Commission submitted its report "Northeast Power Failure", dated 6th December, 1965, to the President.

The origin of the incident was traced to a "back up protective relay". The particular relay setting was reassessed and the critical level at which it was set to function was increased in January 1963. This relay operated normally and isolated an unfaulted line. The power flow on the disconnected line shifted to the remaining four lines each of which became loaded beyond the critical level and tripped.

Quoting from the F.P.C. report, "With the dropping of the lines to Toronto, the power being generated at the Beck Plant and at PASNY'S Niagara plant which had been serving the Canadian loads around Toronto amounting to approximately 1500 MW, reversed, and was superimposed on the lines to the south and east of Niagara. It was this tremendous thrust upon the transmission system in Western New York state which exceeded its capacity and caused it to break up." The consequent chain reaction shut down the system.

It must be realised that apart from the losses incurred by industry and the business community, due to failure of supply, such a castrophic failure may cause damage to equipment. In the North East power outage, failure of the electric pumps supplying lubricating oil caused considerable damage to generator shaft bearings. Consequently the reconnection of the system presented a major problem.

Hydro-Electric stations have a distinct advantage over thermal stations in that they can usually be stated without electric power other than emergency batteries. The time lags in placing thermal stations on load are considerable. Above all it must be appreciated that the co-ordination of the re-energising of such a vast inter-connected system is an immense problem in itself. It must be apparent that the protection and control of large integrated power pools must soon be performed by digital computers. The task is too complex for individuals and the problem involved needs urgent attention.

The Federal Power Commission's report contained nineteen recommendations including the need for auxiliary power sources, stability studies, fully co-ordinated power pools, mixed generation (hydro-electric, pumped storage, coal, etc.), internal load shedding and improved communication facilities.

Following the North East power failure a new philosophy for the analysis and design of large electric power inter-connected systems needs to be evolved. Disaster control procedures providing successive lines of defence, need to be evaluated and implemented. The principles of load shedding, distribution of spinning reserves and system separation must be applied in a more positive manner, with a view to their eventual control by a digital computer which can assess a complex situation and make vital decisions far quicker than a human operator.

18.6 Nuclear Power

It is some eleven years since the first nuclear power plant at Shippingport, Pa, first supplied electric power to the public on a regular basis. Progress from that date was slow, but in 1965 there was a rapid increase in the number of orders for nuclear stations.

It would appear that in the early 1970's about one half of all the new power stations in the United States will be nuclear. Such power projects are already under contract as a construction period of 5 - 6 years is required from the date of authorization. The significant number of recent commitments to construct nuclear plants is clear evidence of their acceptance by power authorities. The acceptance of light water reactors for current contracts has served to speed the progress of advanced reactor projects. The sodium cooled liquid-metal fast breeder reactor (LMFBR) is receiving the highest priority in the development programme of a number of countries.

It is predicted that rapid developments will take place within the next ten years. The accelerated rate of construction of nuclear power stations will give great impetus to the applications of nuclear power to the desalting of sea water, the production of chemicals and metals, the synthesis of ammonia for fertilizers, the elctrolytic production of aluminium and the production of magnesium from sea water by electrolysis.

18.7 Magneto-Hydrodynamic Generators

The conventional alternator has been used for power generation for over 80 years. However, it must be recognized that the effective further development of the steam cycle as used in turbines is nearing an end. Additional gains from increases in temperature and pressure are limited in their economic achievement. Furthermore, the trend during recent years to atomic power sources calls for new and improved forms of power generation. The adoption of a working fluid which becomes an integral part of the electric generation process would be a distinct advancement.

The principle adopted in Magneto-hydrodynamic (MHD) or magneto-plasmadynamic generation is relatively simple. A stream of electricity conducting fluid flows through a transverse magnetic field. As a result an electric field is induced in a direction mutually perpendicular to the flow and the magnetic field and from this induced e.m.f. power may be drawn. Thus the MHD generator is analogous in action to the convention generator. Actually the simple MHD generator produced low voltage direct current and various schemes have been proposed for the porduction of alternating current. Research experiments producing MHD power have benn carried out at the Massachusetts Institute of Technology, Massachusetts, and a number of Universities.To carry this experimentation beyond a certain stage required rather expensive large scale equipment and power prototypes have now been successfully constructed, for example, the AVCO - Everett Research Laboratoty in the U.S.A. has produced a maximum power of 1.5 MW burning oxygen and kerosing/alcohol. The same laboratory has also constructed a Mark V version with an output of 40 MW.

Other research is directed toward determining the potential of the liquid-metal concept for commercial power systems. Binary cycles are being considered such as the coupling of a liquid metal primary cycle with convention steam secondary cycle. The primary cycle would consist of a liquid metal generator tied directly to a liquid metal cooled reactor. It has been suggested that efficiencies of 60% may be feasible.

18.8. Automatic Control

The application of the principles of Automatic Control to items of equipment in power systems is being extended as the systems grow in complexity. Quantities such as voltage, frequency, time control, active and reactive power flow, etc. have been controlled for a good many years. With the trend towards larger boiler and turbo-generator units and the parallel development of instrumentation telemetry, control by on-line analogue digital or hybrid computers is essential to achieve optimum operating conditions. The operations of each power station need to be optimised automatically, in order to achieve the best performance. In fact, large economies can be effected in integrated power systems, by optimising the many variable functions with computer control.

Some of the power system functions which become an essential part of the automatic process are:

1. Load forecasting:
   Load predictions 24 hours in advance.

2. Generator incremental cost:
   The determination of current incremental power costs for generators using fixed data or power plant computation of heat rate for economic despatch.

3. Economic despatch:
   The computation of the most economic loadings for all units in operation based on generator incremental cost curves and transmission losses. It is achieved when the generators within the area are loaded to equal incremental costs of delivered power.

4. Generator selection:
   The selection of the most economic generator units for the increase or decrease in load. Results from dynamic programming achieves the best solution.

5. Interchange accounting:
   The determination of the optimum amounts of power interchange and the statements of accounts for the inter-connected operation.

6. System security and protection:
   A check of overload transmission circuits and relay schedules.

7. Loads frequency control:
   The loading of generators to maintain frequency and interchange flows at desired levels. The settings should be calculated every 5 seconds or less to follow load dynamics.

8. Spinning Reserve:
   The computation of the amount of immediately available reserve necessary to minimize the risk of power shortage while optimizing costs.

Digital Computers are being used extensively for the design, analysis and operation of power systems. All the relevant system design parameters may be stored and software programmes arranged to determine the performance and behaviour of the system. The updating of the system parameters can be carried out by the computer for the addition of new transmission lines, interties, additional power stations, substations, industrial loads, etc.

Furthermore, it must be stressed that with the dynamic growth of already complex integrated power pools there is an urgent need for all operations including protection to be controlled by computers.

Initially we were concerned with the stability of systems having two or three power stations and load centres inter-connected by transmission lines. With the growth of the power network the concepts have changed to such an extent that each power station must be replaced by an "Area" representing in itself a power system. These areas are inter-connected by E.H.V. tie lines; this is the new "stability problem". Each area needs to be controlled within itself and should be responsible for the absorption of its own load changes within its perimeter. This is termed "Area Control". The way in which these areas behave when EHV interties connect them to form an integrated power pool is most important.

We need to study the dynamics of large systems to discover the significant parameters and the properties of the system for control purposes. This all becomes part of the new form of stability problem. The system must remain stable under sudden changes such as the inadvertent loss of component parts under fault conditions.

It should be not imagined that a single computer will control the whole of a power system but we may expect inter-connected computer systems communicating with and controlling each other. Computers located in power plants will report to an "Area Computer" or controller. The Area computers would in turn report to the "integrated power pool" computer. In the future the decisions for control and operation will become far too complex for human controllers. Furthermore these decisions and actions need to be taken quickly to ensure optimum system performance.

Already computer controlled power plants have achieved considerable economics in both power generation and equipment maintenance. The economic savings to be achieved by computer control of a large system would far outweigh the cost of the computer system.

The total investment in the power industry in the next ten years justifies an intensive effort on the part of the new breed of power engineer to solve the problems ahead as the systems grow to the next stage.

18.9 PROBLEMS

Problem 1

Comparing a.c. and d.c. transmission, it has been stated that direct current transmission becomes economical when the savings on transmission lines exceed the higher cost of the terminals. What is the measure of the saving in transmission costs?

Let us assume that the same power W is to be transmitted at the same insulation level; that is, the same peak voltage E to earth is to be adopted in each case. The a.c. line to neutral voltage is $\frac{E}{\sqrt{2}}$ (root mean square value)

Assuming that the line currents are balanced, then the three phase a.c. power is expressed by

$$W = \frac{3E}{\sqrt{2}} I_{ac} \cos\phi \quad \ldots\ldots\ldots\ldots 18.1$$

and for a d.c. line with a voltage ±E above or below earth, (that is, one line is +E above earth and the other –E below earth) the power is

$$W = 2E\, I_{dc} \quad \ldots\ldots\ldots\ldots 18.2$$

If the power factor, $\cos\phi = 1$,

$$\text{then } \frac{3E}{\sqrt{2}} I_{ac} = 2E\, I_{dc} \quad \ldots\ldots\ldots\ldots 18.3$$

$$\text{and } I_{dc} = \frac{3}{2\sqrt{2}} I_{ac} \quad \ldots\ldots\ldots\ldots 18.4$$

Assuming that the same current density will apply for the a.c. line, then the ratio of

$$\frac{\text{d.c. copper per line}}{\text{a.c. copper per line}} = \frac{3}{2\sqrt{2}}$$

and

$$\frac{\text{total d.c. line copper per unit length}}{\text{total a.c. line copper per unit length}} =$$

$$\frac{3}{2\sqrt{2}} \times \frac{2}{3} = \frac{1}{\sqrt{2}} = .707$$

Apart from the saving in conductors for d.c. transmission, there will be a saving in insulation and transmission line tower costs.

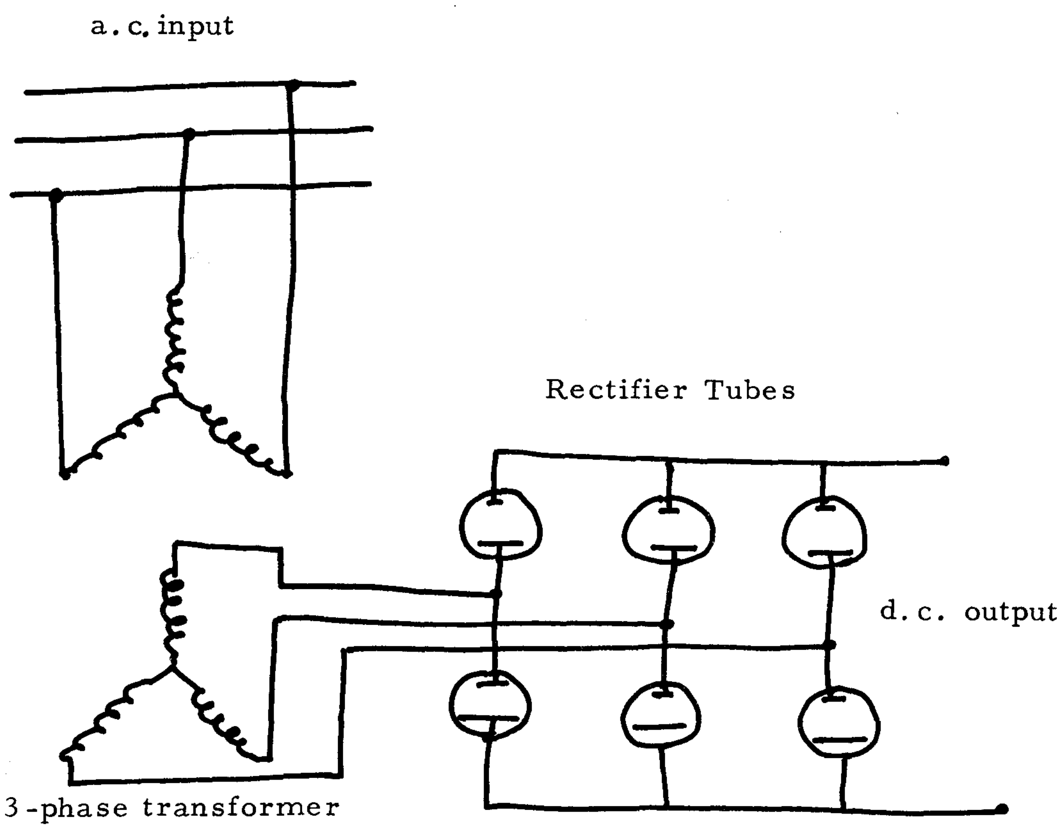


Figure 18.1 A Six pulse two-way converter

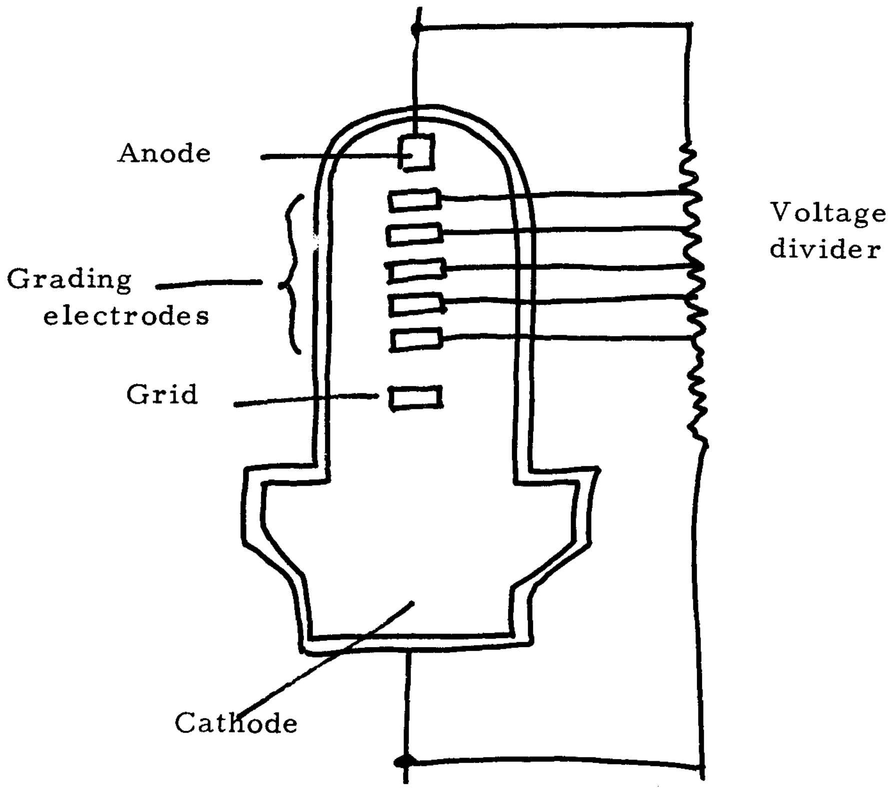


Figure 18.2 A High Voltage Valve

18.2 **Problem 2**

How would you arrange the flow of water in a tidal power scheme?

Methods

1. Single way operation on emptying

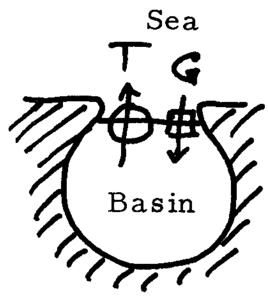


Figure 18.3

2. Single way operation filling.

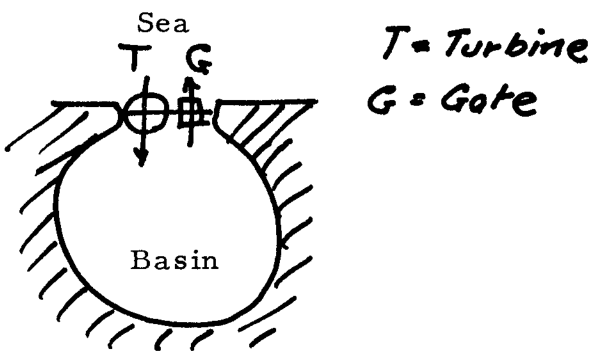
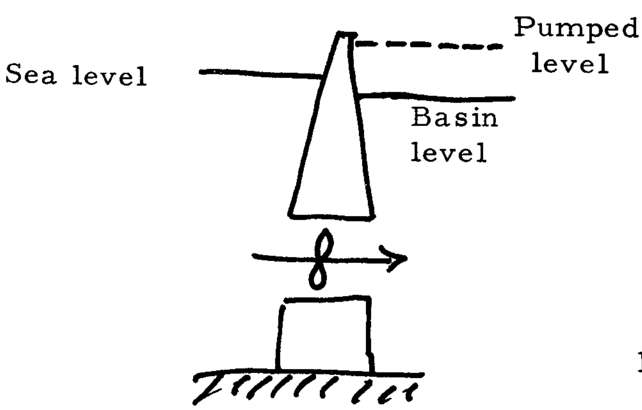


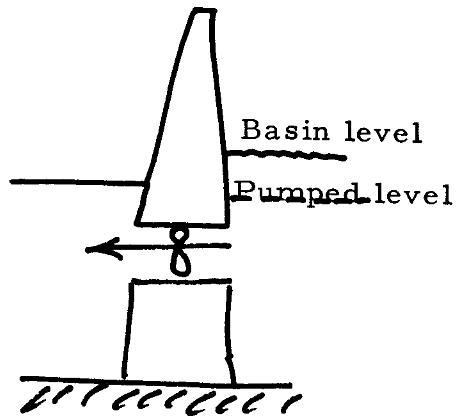
Figure 18.4

Both methods may be improved by pumping (a) at high tide (b) at low tide. In the first case of pumping at high tide, the level of the basin can be raised when the level of the sea and the basin are approximately equal. Relatively little energy is needed to raise the water level but when the tide has fallen this 'pumped' water falls through a greater head.



Pumping high tide

Figure 18.5



Pumping low tide

Figure 18.6

4. Double way operation

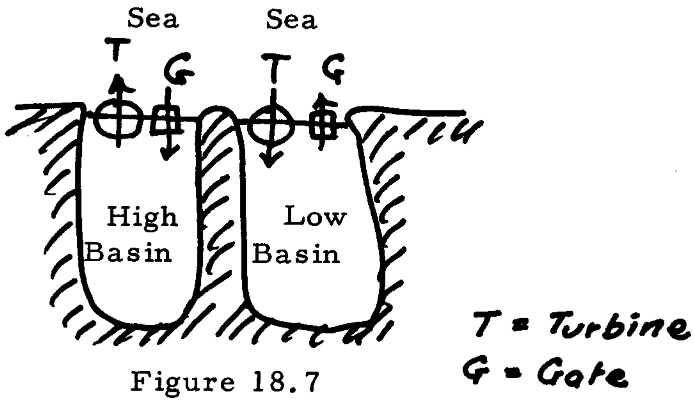


Figure 18.7

5. Sesqui operation

This cycle makes use of two successive spring tides to pass the water three times through the turbines, with two intermediate pumpings

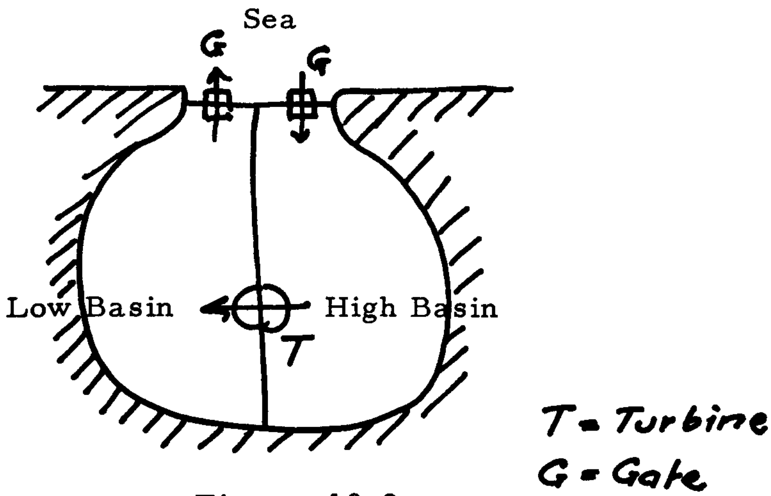


Figure 18.8

Many other schemes and refinements could be proposed.

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Chapter 19

HUMAN SYSTEMS

"We have an innate propensity to get ourselves noticed, and noticed favourably, by our kind. No more fiendish punishment could be devised, were such a thing physically possible, than that one should be turned loose in society and remain absolutely unnoticed by all members thereof."

William James 'Psychology' Ch. 12.

Chapter 19

HUMAN SYSTEMS

(With reference to modern concepts of management)

19.1 Introduction

Science has already become a most persuasive force in our modern society, having a widespread influence over all of man's activities. It has had an impact on man's work environment, on his standard of living, on his health, on his social institutions and on his recreational activities.

Scientific and technological achievements have had a similar impact upon business organisations. Almost every industry has seen major advancements not only in its products but also in its means of production. Scientific and technological advancements have greatly influenced decisions made by governments on such matters as health, education, foreign policy, welfare and national defence.

There are many other areas where science and technology will undoubtedly make spectacular strides in the years ahead – in biological and medical research, in nuclear technology, in computers, in transportation and communications systems and in space technology; and it would appear that this advancement will continue at an ever-increasing rate.

As major scientific and engineering projects become more complex, the problems of managing these projects become more difficult. Scientific and technological advancement requires a high degree of specialisation among men and organisational units. These two factors, extreme degree of specialisation and the emergence of large and complex systems produce far reaching human problems, some of which are as follows:-

First: There is the need for someone to translate the scientific achievements into beneficial and usable products or services: and the achievement of human, moral and ethical controls over the utilization of many of these developments (e.g. nuclear energy) is as important to mankind as the discoveries themselves.

Second: There is the problem of optimum allocation of resources. Many of these developments will present simultaneous claims on the limited material, human, natural, financial and time resources. The means (resources) available at any point in time will never be sufficient to meet our wants (objectives), and choices will have to be made. Criteria for determining priorities and for evaluating the proposals will present a formidable task.

Third: Accelerating technology leads to shorter and shorter lifespans for each new generation of products (e.g. computers). Planned obsolescence has become a byeword in modern industrial society. In addition the amount of money and time required to design and develop the product and set up production facilities has increased. This longer development cycle has made the management function of long-range planning mandatory in order to minimise the risk of expending valuable resources on a product with little chance of success.

Fourth: Large and complex programmes require for their solution greater specialisation in narrower fields. As a consequence the management of such programmes requires the co-ordination of an enormous number of people – many of whom will have specialised training in a great variety of fields. With more specialisation, integrating the activities of the people into an effective organisational unit will be increasingly difficult.

Fifth: The problem of conflict between managers and scientists. The managers major drive is to obtain a reasonable return on the company's investment within a reasonable time and within the patience of his board of directors and shareholders. The scientist on the other hand wants to expand the limits of his knowledge and to apply this knowledge through an organisational team to stimulate this team to get results, and to get the necessary tools from the organisation. Conflicts arise when the scientist wants tools and equipment which the project cannot afford. A great deal of attention will have to be given in the future to studies of the factors motivating scientists, managers and other employees if optimum results are to be achieved.

Sixth: With the emergence of very large and complex projects it will be necessary to perform managerial functions over a very broad spectrum of intra company, inter company and inter industry relationships. This will require new and more effective approaches to the management of these programmes.

Seventh: The development of large and complex projects also introduces uncertainty into the decision making structure. Uncertainty may arise in respect to the selection of a particular project and this will be met by the development of better techniques for evaluation. Another area of uncertainty relates to costs. Large projects require large amounts of money – but it is often not possible to define in advance and within reasonable limits the amount of money likely to be involved. (For example, the original amount of money requested for the development of the atom bomb in America in 1942 was $6 million – within six months the estimate was raised to $400 million.) This area of uncertainty poses economic problems of far reaching importance. The impact of such uncertainty on the internal and external balance of the economy, the effects of changes in the distribution of income and money flows, and the impact on the members and institutions of the economy are matters of extreme importance.

These are some of the problems now emerging. Technological and scientific developments will continue – but whether or not these developments will benefit mankind as it is hoped, depends on the ability of man to work together as a team. He must consider well the role he is playing, he must consider not only the immediate and short term effects of his actions but also the longer term implications of these actions on the environment and on his fellow men.

In short, the solution of these problems cannot be left to one section of the community alone be they scientists, technologists, politicians or managers – all must share the responsibility in determining the shape of tomorrow. The development of effective and efficient "human systems" is more important now than it has ever been in the history of mankind. In this respect management has a unique opportunity to contribute to a more effective comprehension of the problems involved.

19.2 Management Defined

The word 'management' is normally used in a very loose fashion. To many, the process of managing is seen as "decision making". But the sort of decisions that have to be made in business are usually ones of engineering, of marketing, of production, of finance and so on. As a result there is a tendency to assume that the study of engineering, marketing, production, finance and so on are part of the study of management, which they are not; these disciplines are the tools of management. Undoubtedly a knowledge of engineering or finance or marketing or production is essential to most firms, but one can have in a firm engineers or accountants or marketing experts who do not necessarily manage – they may be advisers. One can have a head of a department of engineering who is an engineer; similarly one can have managers for several different departments each of whom is interested in the subject specific to that department whether it be sales or procurement or personnel. What is common to all of these departments is the fact that managers are men who manage people. Thus management is concerned not specifically with marketing, finance, engineering and production - - these are the "tools of management" – but is concerned with planning, commanding, controlling, co-ordinating and organising. Tool subjects such as finance, engineering and marketing are used in order to come to the conclusions and decisions that are necessary for action in management.

To define management in this way helps to focus attention on the fundamental administrative processes which are essential if any organisation is to meet its primary objective. These basic managerial processes are required for any type of organisation – business, government or other activity where physical and human resources are combined to meet certain objectives. Furthermore these processes are necessary regardless of the specialised area of management – production, distribution, finance, etc.

Koontz and O'Donnell 19/1 suggest the following five basic functions –

1. *Planning* – The function of selecting the enterprise objectives and the policies, programmes and procedures for achieving them. Planning is, of course, decision making since it involves choosing among alternatives.

2. *Organising* – The function of determining and enumerating the activities required to achieve the objectives of the enterprise; the grouping of these activities; the assignment of such groups of activities to a department headed by a manager; and the delegation of authority to carry them out.

3. *Staffing* – The function of staffing comprises those activities essential in manning and in keeping manned, the positions provided for by the organisation structure.

4. *Direction* – The executive function of directors embraces those activities which are related to guiding and supervising subordinates.

5. *Control* – This function includes those activities which are designed to compel events to conform to plans. It is thus the measurement and correction of activities of subordinates to assure the accomplishment of plans.

Although these five functions are listed and described separately they should not be considered as independent activities; nor is any exact time sequence implied. Adequate performance of any one of these functions is dependent upon the performance of the other four.

These activities have been and will continue to be fundamental functions in the management process. If improvements are to be sought in the managerial process to meet the needs of this advancing technological age it will be necessary to re-examine the assumptions upon which the present theory of management has been based. These assumptions, regarded as plausible under earlier conditions, may well have to be changed in the light of new knowledge and new conditions.

Before proceeding it will be useful to indicate, briefly, the influence of changes in scientific thought on assumptions basic to management theory.

19.3 Science and Management.

From the very beginning of time man has involved himself continuously in efforts to understand nature and to adapt his physical environment to his own benefit. Man has also devoted his efforts to understand himself and to develop social

organisations as frameworks for group activities – but the development in his knowledge of the physical environment has outstripped his ability to develop human relationships necessary to make full use of his technical achievements. For many centuries Aristotelian science dominated western thought with the belief that environmental problems could be solved by reasoning from self-evident principles. The goal of this reasoning was to explain 'why' things happened.

Modern science, with its origins in the Renaissance places its emphasis on "how" things happen. Galileo emphasised the method of controlled experiment, which has since formed the basis of all scientific investigation. Newton carried this method forward in his endeavours to find order in the universe. He used this method to explain many of the processes of nature under a mechanical model with well defined cause-and-effect relationships. However, even though a model may adequately represent the system under study when it is constructed, very few models are designed to cope with a dynamic environment.

As new phenomena are discovered or changes in environment are experienced old models have to be reviewed. Such was the case in the beginning of this century when it was found that new phenomena such as x-rays, gamma rays, radiation and sub-atomic particles, could not be explained by reference to the Newtonian system. The possibility of negative gravity, the discovery of increasing numbers of sub-atomic particles and the detection of anti-particles led to the extension of Newtonian laws and to the development of quantum mechanics. An important outcome of quantum mechanics was Heisenberg's principle of uncertainty. This principle states that as certainty about the position of a particle becomes greater, the greater is the uncertainty about its velocity and vice versa. Perhaps what is more important than the discovery of this new knowledge is the fact that it is now realised that cause-and-effect relationships do not exist in all physical situations.

One of the most important differences between the physical sciences and the social sciences is in the outstanding mathematical machinery which the physical sciences are able to employ. Even so mathematical machinery alone is not sufficient to make deductions in an atomic age.

Social science has also experienced profound changes in the twentieth century – and these changes in many ways reflect the quantum mechanical and relativistic developments in the physical sciences.

Many of the traditional economic theories relied on the assumption of order and certainty. It was assumed that the events of the world could be predicted by knowing the structure of this order. In a similar way it was believed that since the behaviour of individual firms and industries was determined by the interaction of impersonal market forces (such as the forces of supply and demand, the marginal rates of substitution, etc.) there was no need to deal with the human side of the enterprise, such as the decision making process, the problems of organisation and control. Thus the economic theory of the firm emphasised economic goals, and objectives, i.e. profit and/or wealth maximisation rather than the psychological oriented goals, i.e. satisfaction of expectations and aspirations. However, the realisation of the inadequacy of such theories to explain the real world led to the development of a new theory called the 'behavioural theory'. The behavioural theory of the firm is a reflection of both the quantum mechanical and the relativistic revolutions in the physical sciences. The focus is now directed to determining why goals are set, why decisions are made, why decisions are implemented. It is now fully recognised that further scientific and technological advances are very much dependent on our ability to develop more effective managerial and organisational abilities.

Whilst substantial progress has been made in management analysis and theory, this progress falls short of meeting the needs of organising, planning and controlling the highly complex technological and scientific undertakings which are now emerging.

The newer approaches to defence and other large scale complexes which are mentioned later in this chapter involve questions of economics, of politics, of psychology, of science, of technology as well as military factors. This new approach recognises the need of newer and more appropriate managerial arrangements in which participation and team work will be required among people experienced and academically trained in many disciplines.

The question is – how is such managerial theory and skill developed?

Ever since man got together to work in groups for a common purpose such as survival there has been management – there has been planning, co-ordinating and ordering. But the process has become more and more complicated as man's civilisation has evolved. To use the analogy of bridge building – man first learned to cross a river by the primitive form of a fallen tree trunk. He then went on to develop new methods of crossing such obstacles eventually learning to build bridges in stone. He learned to do this by trial and error; and from his experience in putting stones together certain principles emerged; for instance, that an arch is one of the strongest structures and requires a key stone at the top to give it fundamental strength. The Romans built beautiful bridges but they had no theory of strength of materials, of friction, of co-efficient of expansion or the like; and we still have people who can build bridges without any theoretical knowledge guided only by intuition derived from practical knowledge based upon experience and observation.

Today, however, we require more than these simple though effective methods of bridge building. We have to build longer, higher and stronger bridges. We have to take account of costs. The Romans were not constrained to think in terms of costs or labour or the effects on the human and institutional system in which they operated. The fact that Roman bridges have

lasted over the ages may indicate that they were overdesigned. It points out the need for concern on the part of designers of to-day and to-morrow to design and construct projects so that the quality and precision is related to the purpose for which the projects are designed. More quality or precision than is required to meet the needs is costly and wasteful of economic resources. Less quality and precision is equally bad as it results in inefficiency and loss. When engineering and other technological projects involve costs which represent a substantial part of a country's national expenditure, the consequences of irrational or inadequately managed programmes can have damaging effects of great magnitude.

So to-day something more than 'practical' management is required. What is required is a stronger, more economic and efficient form of management in order to handle complex problems in our rapidly – changing economy. New theories are required to build these new forms of management.

But even though a person understands the theory of management, this knowledge does not necessarily make him a manager. An engineer may study theory and make calculations but he cannot become an engineer unless he has practical experience. He must know just what theory can do when applied and this he has to learn by experience of theory in practice. Theory is not enough, there must be some knowledge of practice; equally experience is not enough, there must be some knowledge of theory.

The vast majority of managers have learned management by precept and by experience. But many managers find that when they move from one situation to another they have to spend a considerable time learning about the new situation for, so often the detailed operational principles learned in one situation are not applicable to another.

Whereas theoretical principles must, by nature of theory, be universal and capable of application in any situation. Therefore, if a man knows theory and principles stemming from it he can quickly adapt himself to any new situation.

What can we say about the Theory of Management?

19.4 Towards a Theory of Management

The test of any theory is its capacity to explain and predict. However, theories are based upon certain basic statements which are assumptions regarded as plausible for the universe of discourse, i.e. the subject under discussion. These assumptions are based upon experience or tests which are carried out or they may be based on "intuition". But "intuition" will not help unless it derives from knowledge.

The Newtonian Physics explained many of the phenomena of the physical world when examined within the framework of existing knowledge. However, the acquisition of new knowledge of physical phenomena required an extension of the Newtonian physics in order to explain the problems which have been apparent from the new knowledge. So it is with any theory. As new knowledge is acquired the assumptions and basic statements upon which existing theories are based must be extended.

Many extensions have been made to the theory of management. Enough is known about the theory of management to explain most phenomena but not enough is known to be able to predict what is needed. In fact it is not at all certain that management knows what is needed. The reason is simple, managers are concenred with managing men, and since men are all unique the number of variables to be considered is fantastically large. Nevertheless there are theoretical premises which do assist us to predict though we remain aware of their limitations.

Some people believe that a theory of management must be concerned with subjects such as economics, finance, marketing, production and so on; and that it is necessary to study these alone to understand management. There are others also who believe that, since getting action implies co-ordinating men, any theory of management must be sociological or psychological. But as already stated, finance, marketing, engineering, etc. are the tools of management and management stands by itself as a theoretical discipline. However, when it comes to the problem of operational principles they cannot be based on a theory of Management alone – other theories, concerned with the tools of management, may have to be brought into such principles.

This has led to some confusion in the formulation of a theory of management. Adam Smith (in his "Wealth of Nations") formulated the idea of division of labour. Taylor (in his "Scientific Management") developed the concept of the scalar chain, the hierarchy of authorities – the idea of 'line and staff'. The structure was a system and pattern to achieve an objective, and believed in the idea of a limited span of control (no one person should supervise directly more than five other persons). There was little attempt to incorporate the theories of the behavioural sciences – man was regarded as an efficient but unpredicable machine governed by purely economic motives. The failure of theories based upon these limited ideas led to the development of analytical studies involving the use of models.

19.5 The Use of Models in Management Studies.

Models represent a means of analysis to evaluate alternative courses of action and to provide a basis for executive action. A model is a representation of reality by its main properties and the essentials of reality are expressed in symbolic form.

Models can be classified by distinguishing physical, schematic and mathematical types.

Physical models are the most easily understood since they look like the object under consideration. Examples of physical models are those used to demonstrate shape and orientation of continents. A model of the solar system is used to demonstrate the orientation of the sun and the planets. Some physical models are used in the simulation process. An aeronautical engineer may test a specific tail assembly design with a model aeroplane in a wind tunnel.

Schematic models are obtained by taking a state or event and reducing it to a chart or diagram. The schematic model may not look like the world it represents. The organisation chart is a schematic model. It is a representation of the state of formal relationships existing between members of the organisation. The value of a schematic model lies in its ability to describe the essential aspects of the existing situation. It does not include all the relationships but concentrates on a single facet. Thus the schematic model is not a solution but only facilitates a solution. After the model has been carefully analysed, a proposed solution can be defined, tested and implemented.

Mathematical models employ the language of mathematics and like other models, may be a description and then an explanation of the system it represents. Mathematical models provide a much higher degree of abstraction and precision in their application. Almost all mathematical models are used either to predict or control.

Mathematical models directed to the study of operations differ from those traditionally used in the physical sciences in two important ways. First, since the system being studied usually involves social and economic factors, these models often incorporate probabilistic elements to explain random behaviour. Second, mathematical models formulated to explain operational systems incorporate two classes of variables – those under the control of the decision maker and those not directly under his control. The objective is then to select the variables under his control so that some measure of effectiveness is optimised.

All of these types of models are employed in the analytical studies of the management tool subjects such as finance, marketing, production and so on. However, the type of model best suited to the study of management organisation is the schematic type.

Many models of organisational behaviour have appeared over the last hundred years. In reading the following brief outline of the development of organisational models it is well to keep in mind that like theories, models are based upon assumptions of reality. As knowledge of reality is increased models and theories must be extended and revised. Models or theories are only capable of explaining relationships in terms of the framework of that model or theory. This point can be emphasised by reference to the application of sampling statistics. In statistics, a universe refers to all of people or firms or materials or items that one wishes to study. In order to reach these items in the universe a statistician uses a frame – which may consist of a list of names, a file of cards, segments of areas or a set of rules. When a sample is selected, it is selected from this frame and inferences from the sample are valid in respect to the frame but may not be valid for the universe. This is so, because there may be gaps between the frame and the universe – i.e. where the frame fails to cover the universe.

Similarly with theories and models, inferences and predictions may not be valid for the universe. As new knowledge is gained or new phenomena is discovered the frame must be extended and the theories revised.

Models of organisational behaviour which have developed seem to be sequentially related to man's psychological and secondary needs. The hierarchy of these needs may be classified as first – his subsistence needs; second – his needs of social and economic security; third – his need for affiliation and recognition and fourth – his need for self-fulfilment. As man's standard of living, education and social conditions improve so he seeks more advanced need satisfactions. The model of organisational behaviour that was developed to meet the lower order need structure (subsistence needs) has been called the 'autocratic' model. This model prevailed in the early part of the Industrial Revolution. Under autocratic conditions the organisation is seen as a relationship between a boss who gives orders and the employee who obeys the instructions. The employee is dependent on the boss.

Although modern observers have a tendency to condemn the autocratic model of organisational behaviour, it is in fact a useful way to accomplish work. It was successfully applied by the empire builders of the 1800's. It helped to build great railroad systems in the United States, and to operate giant steel mills, and produce a dynamic industrial civilisation in the early 1900's. It was, no doubt, a reasonably effective way of management, when there was a 'benevolent autocrat' who had a genuine interest in his employees, and the employees accepted the fact that their role was one of obedience to the boss's orders. Here then was one conceptual model of an organisation. The question is, of course, will this autocratic set-up lead to maximum utilization of human resources? Will this type of organisation achieve its goals efficiently, speedily and economically, and encourage maximum effort? This conceptual model still exists in some organisations and from the observations made it has been found to be ineffective. A compliant employee might not always talk back to his boss but he certainly "thinks back". It has been found that, under this type of organisation employees possess a seething mass of insecurities, frustration, and aggressions towards the boss. The question is – How does this sort of relationship affect the motivation to work?

It seemed rather obvious to progressive employers that there ought to be some way to develop employee satisfactions and adjustments during production – i.e. some way in which individual goals of employees may be fulfilled. If the employee's

insecurities, frustrations and aggressions could be dispelled, he might work more effectively. It was found that the employees were frustrated with the work conditions, some feared insecurity from an income point of view, and some feared the insecurity of the employment conditions. It was soon recognised that individuals possess personal goals and that if ways and means could be found to satisfy those personal goals then employees may be more constrained to work most effectively for the principal organisational goals.

So an improved model was developed. One such model was called the custodial model. The custodial model originally developed in the form of employee welfare programmes. During the depression of the 1930's emphasis changed to economic and social security. During and after World War II the main focus was on specific fringe benefits as a means of higher motivation. A successful custodial approach, however, depends on economic resources and an organisation must have economic wealth to provide this sort of economic security, pensions and other fringe benefits to its employees. However, experience with this model of organisation has shown that over time the employee became preoccupied with maintaining his security and benefits, rather than with production. It has been found that these benefits do not strongly motivate an employee. In fact, the custodial model leads to a condition of dependency by the employee on the organisation. As a result he did not produce much more vigorously than under the old autocratic approach. Security and contentment are certainly necessary for a person but they are not themselves very strong motivators. The result was a passive co-operation by employees. They were pleased to have the security but as the employee grows psychologically, they also sought more challenge and autonomy.

That man was not a machine and is activated by more than economic motives was first recognised by the Industrial Health Research Board in the First World War, and formed the basis of the "Human Relations School" in the 1930's. The existence of informal social groups and the conflict between personal goals and those of the enterprise became a matter of prime importance. These studies revealed that the division of labour led to specialisation, isolation of people, a sense of insignificance and a lowering of the motivation to work and so to a decrease in productivity.

As a result of further research in behavioural science, an improved model of organisation has been formed and it is called the supportive model. The supportive model establishes a manager in the primary role of psychological support of his employees, rather than in a primary role of economic support (as in the custodial model) or the "power over employees" (as in the autocratic model).

The supportive model depends upon leadership of power or economic resources. Through leadership management provides a behavioural climate to help each employee grow and accomplish, in the interests of the organisation, the things of which he is capable. The leader assumes that workers are not, by nature, passive and resistant to organisational needs but that they are made so, by an inadequate supportive climate to work. They will take responsibility, develop a drive to contribute, and improve themselves, if management will give them half a chance. Management's orientation, therefore, is to support the employees performance. Now it has been emphasisedthat the concept of motivation is indispensable for understanding the working of the model. Unless the employee has a genuine interest in accomplishment he will not contribute effectively to the organisation. The organisation may set up attractive retirement systems, and programmes for recreation, safety and health. They may keep the employee happy on the job, but unless strong motivations are built up towards the work itself, reinforced by attitudes towards the superior, and towards fellow employees, high productivity will not necessarily follow.

Attention to the factors affecting motivation are of special importance, when considering a model of an organisation employing highly specialised and professional personnel such as people involved in the scientific and technological fields.

The model that is conceived of this sort of organisation should suggest the question – under what sort of conditions are strong and weak research motivation found? Should a research leader direct his staff – i.e. should he assume personal direction of all work and treat his staff, as so many extra hands to execute his ideas. Or should he adopt a passive role and assume that his scientists know what they are doing and only help when asked, or should he actively participate in a team. Other questions may be asked – what motivates a scientist? – is it his scientific interest? – the prestige of working with other scientists? – the role played by the chief scientist? – or what?

The sort of model formed to examine these factors has been called the collegial model. The collegial model is still evolving. It has developed from recent behavioural science research in motivation factors and the work of a number of people in project management and matrix organisations. The collegial model readily adapts to the flexible intellectual environment of scientific and professional organisations, working in substantially unprogrammed activities, which require effective teamwork, scientific and professional employees who desire the autonomy which a collegial model permits and they respond to it well.

The collegial model depends on managements building a feeling of mutual contribution among participants in the organisation. Each employee feels that he is contributing something worthwhile and is needed and wanted. He feels that management and others are similarly contributing so he accepts and respects their roles in the organisation. Managers are seen as joint contributors rather than bosses.

The managerial orientation is towards teamwork which will provide an integration of all contributions. Management becomes more of an integrating power than a commanding power. The employee response to this situation is responsibility. Feeling responsible, the employee disciplines himself for team performance in the same way that a football team member, disciplines himself in training, and in game performance.

Because of the development of very large and complex projects, such as the weapons and satellite projects carried out in the United States and elsewhere, it has been necessary to establish new and more effective approaches to the management of these programmes.

19.6 **Programme Management Concept.**

One such development is the 'programme management concept'. The programme management concept is a dynamic philosophy geared to changing managerial requirements in the research development, procurement and utilization of large scale military and civilian systems. Basically it is the philosophy of the integration of all the activities necessary to the accomplishment of the primary goals and objectives. It is based upon the integrative and co-ordinative activities of focusing a myriad of functions on the accomplishment of total organisational goals.

The first step in the evolution of the programme management concept was the product mission concept. Any product, civilian or military, can be thought of as having a mission to perform. In order to fulfil the product mission it is necessary to perceive a need, to design or develop a successful product, to produce it efficiently, to stimulate demand for this product, to provide distribution through channels necessary to reach the consumer and perhaps to educate potential users in proper utilization of the product. The accomplishment of the successful mission for the product can be viewed as a programme management problem much broader than just physical production. A product mission is measured in terms of ability to optimize all the functions necessary for its success, rather than just to maximise the efficiency of performance of any one function.

Although there are many complex civilian examples of the necessity for programme management some of the most complex examples are seen in meeting the product mission of many of the new weapon and space systems of the United States. The size and complexity of these projects has led to a new concept of organisation and a new concept of the organising function.

The organisation function means the integration of people and other resources into a system so that the act will lead to the accomplishment of predetermined goals. Many activities are involved in the organising function – the determination of activities required, the allocation of authority and responsibility in the establishment of control mechanisms to measure the adequacy of performance. The new concept of an organisation is thus, "an assemblage of materials, machines, technical and scientific facilities and personnel geared to task accomplishment through series of interactions and integrated into a total social system". The variety of organisations are numerous and include non-military government agencies, military departments and private industrial firms. In spite of the fact that there are different types of organisations, there appear to be certain characteristics common to all of them, these are as follows.

1. Growing Size and Complexity. The growing size and complexity of projects has created many problems in integrating the various segments or parts; also, the task of integrating activity on a number of varied programmes into effective total organisational performance has become more and more difficult.

2. Specialisation of Skills. Advances in science and technology have led to a demand for further specialisation of skills. This is true for the skills of the scientist, the technician and even the manager. With more specialisation, integrating the activities of the people into an effective organisational unit is increasingly difficult.

3. Diversity of Objectives. Many organisations have a diversity of organisational and personal objectives. There is the possible conflict between the manager and the scientist within the organisation in such spheres as budgets, costs, design changes and delivery schedules. One of the major managerial problems is that of integrating these diverse and conflicting objectives into the total organisational performance.

4. Adapting to a Changing Environment. There is a growing requirement for organisations to adapt to changes resulting from the advance in science and technology, and from other factors in their environment. It is impossible to think of these organisational units as static systems. They have to be dynamic and adaptive in order to meet the challenge of continued change.

With the advent of the newer more complex programmes both military services and private companies have had to adapt their organisational structures to augment traditional functional departments. Pressures of technological complexities and time requirements have led, in some cases, to the setting up of centralized management agencies whose primary responsibility is to provide the overall integration on a system basis which is lacking in a strictly functional organisation. The traditional organisational relationships have been altered to meet the requirements of multi programme operations. As organisations have had to integrate efforts on a number of major programmes the traditional approaches have become less effective. Thus there has emerged the concept of programme managers, whose primary responsibility is the performance of all functions necessary for successful mission accomplishment.

In spite of the evolution of the project organisation, the matrix type organisation is still considered to be important. Basically the concept of the matrix organisation is one wherein there are well established functional departments which have special skills and capabilities for performance of these functions for a variety of programmes. Essentially the programmes flow through the functional complex and receive the services of these specialised departments. Furthermore, the matrix organisation includes the idea of a number of programmes being worked on by the functional organisation simultaneously.

19.7 Conflict between Programme and Organisation.

One of the problems that has arisen concerns management where they are operating primarily to maximise a specific programme where such an approach may conflict with total organisational goals. It has been stated that there are many aspects to the problem of possible conflict between optimal programme performance and overall organisational objectives.

First, if a company is to continue in business they will have to execute their programme with efficiency and expertise. However, they will endeavour to organise in such a way that will not affect their overall capability. A slavish adherence to either a completely matrix or a completely projectised way of doing business can lead to undesirable consequences. A company which adopts a matrix organisation retains the functional skills involved, and this enables it to apply across all projects, the knowledge gained on one. On the other hand, it does not motivate those managers who lack a broad perspective. The adoption of a project organisation focuses the talents at hand on the job to be done, everyone knows his job, who his boss is, and what responsibilities he has. It helps to determine what is required for various operations and it helps to develop people with capacity for systems development, yet there are difficulties in the project typed organisation. It makes cross fertilization of ideas more difficult, it makes the transfer of people difficult, and as projects are completed employees often become uneasy. This has been evident in the Snowy Hydro-electric Scheme.

It has been found to be possible to solve many of these problems by changing the type of organisation as a function of the phase of the programme. In the early phase management should apply the best talents available across the organisation and matrix organisation is generally best, then when the programme definition phase is past and the goals become clear, the project organisation is more effective and finally, as the project phases out they should go back to the matrix operation. However, finding the right organisational pattern is a matter of continued adjustment, it is not always clear that what is best for the programme is best for the long run goals of the organisation.

Appraisal and Control of the Programme: One of the problems has been found to be that of controlling the performance of too many objectives which often were in conflict. For example, meeting both time and cost objectives in advanced technological programmes may be incompatible. It has been suggested that controlling agencies often have the benefit of hindsight appraising the performance of a particular programme and this did not give full recognition to the great difficulties involved in forward planning in major research and development programmes. Also it has been suggested that rigid control in the early research phases could lead to restrictions and to reluctance to move ahead on new ideas and experiments.

19.8 Summary and Conclusion.

It is apparent from the literature that many of the newer concepts most effective in organising for goal accomplishments are coming from the advanced technology industries. They are innovating with respect to managerial processes and organisational relationships and have developed approaches significantly different from traditional concepts. The organisational changes taking place in the advanced technology industries eventually will be employed in other more stable industries. Traditional organisational concepts are sufficient when we have a relatively quiescent environment, moderately advanced technology, and clearly defined and easily accomplished objectives.

When an organisation is dealing with large and complex programmes requiring the integration of science and technology and demanding more effective integration of sub-systems, it is found that the traditional concepts of management organisation are no longer appropriate. Many of the traditional concepts of line and staff relationships, hierarchical structures, span of control, and other traditional organisational concepts, do not meet the demands of the advanced technology programmes. In the future the problems of organising for effective performance will require the greatest level of co-operation and understanding by scientists, technologists, management and government. The growing need for better means of co-ordination and integration will be increasingly crucial with increased scientific and technological complexity and larger scale programmes. One of management's major responsibilities in the future will be the organisation and efficient utilization of human and physical resources for mission accomplishment.

Conflict of Interest between Managers and Scientists. Many of the conflicts that arise or are likely to arise between scientists and managers stem from their respective training and motivations.

Most engineers are motivated by the need for achievement and recognition in their individual scientific discipline. This is not to underestimate their needs for economic and other job benefits.

On the other hand recognition must be given to the factors which motivate the manager. Organisations will have increasing problems of providing reward systems to motivate scientific personnel. Most organisations, particularly in business, have been set up on a hierarchical basis to provide a reward for the manager moving up to the higher and higher positions in the organisation. The new situation may demand the establishment of a dual hierarchy wherein the scientist can move up even though he does not assume managerial responsibility. The question of whether or not the scientist should be required to move out of the laboratory and into the management position in order to receive major organisational rewards of money and status has often been discussed. There seems to be general feeling that organisations should provide a reward system to remunerate the scientist for his pursuit of scientific and technological knowledge.

In the light of the far reaching consequences of scientific developments a question ı .ay well be asked whether or not engineers should be given the opportunity to share in the responsibilities of managerial decision making.

Perhaps this conflict of interest suggests the need for more effective communication between scientist and manager. Perhaps the communication should be fostered at the student level so that the implications of precise mathematical formulas of the physical sciences and lack of precision of the theories of the "dull science" may be more fully comprehended by both classes of students.

The Role of Management: The role of management has been stressed as the most critical factor in pushing advanced technology programmes through to meet their goals. It has been suggested that the manager's primary role was one of integrating conflicting interest groups towards specific goal accomplishment. The point has been made that management should be charged with the responsibility for adapting the organisational resources to changing environments in order to maintain scientific and technological progress.

Perhaps a thought should be spared for the development of techniques to enable management to give consideration to the long term effects of "goals", "projects" and "missions" on the community, on the nation and on the general environment. A question might well be asked whether or not sufficient attention is paid to the long term effects of policies and decisions.

Systems Management Concepts: The growing complexities and diversities of operation in today's advanced technology programmes makes it necessary to conceptualise a system basis in order to integrate these complex operations. Primarily a systems concept is a way of thinking about the job of management which provides a framework for visualizing the internal and external factors affecting the organisation as an integrated whole. It provides for recognition of the proper role of functions in sub-systems. Systems are thought of as an organised or complex whole, assemblage or combination of things or parts forming a complex or unity whole. The working systems concept which has become increasingly important in military management is an example of the application of this systems concept. Management is the primary force within the organisation which co-ordinates the activities of the sub-systems and relates them to the environment. Essentially management is the process of relating human and material resources into a total system for objective accomplishment.

Changing Management Concepts: It has been pointed out that the primary reasons for the emergence and application of the systems concept has been advancing technologies and increasing industrial complexities within society. These forces will continue possibly accelerating in the future. The systems concept will allow more effective adaptation to scientific and technological environments. While this approach has been, and is currently being implemented in most of the more advanced technology industries, its use will spread to other industries in the future. One of the major changes within business organisations will be the breakdown of functional specialisations geared to optimizing performance of particular departments. There will be growing use of organisational structures designed around projects and information decision systems. The systems concept calls for the integration into a separate organisational system of all those activities which are related to particular projects or programmes. Business organisation must be considered as a sub-system of a larger environmental system. Even in industry or inter-industry, systems must be reorganised as sub-elements of an economic system and the economic system should be regarded as part of the society in general. The point that must be emphasised is that the system concept is primarily a way of thinking, a mental frame of reference which can be utilized by management in performing its traditional primary functions of planning, organising, and controlling operations. These activities have been and will continue to be fundamental functions in management process. The systems concept provides a new framework for carrying out the integration of these activities. The newer patterns of management and organisation which will have a broad impact upon other industries and agencies in the future.

19.9 Exercises

1. Define science, technology and management.
2. What are some of the problems emerging from the advances in science and technology?
3. What is meant by a programme management concept?
4. Suggest how the conflicting objectives of the scientist and the manager can be compromised.

19.10 Suggested Reading

Koontz and O'Donnel – "Principles of Management" (McGraw Hill, 1959)

Cohen and Cyert – "Theory of the Firm" (Prentice Hall, 1965)

Kast and Rosenweig – "Science Technology and Management" (McGraw Hill, 1962)

Fabrycky & Torgersen – "Operations Economy" (Prentice Hall, 1966)

Chapter 20

LEARNING, ADAPTIVE AND GOAL SEEKING SYSTEMS

"This is an approximate idea only. Our maps
always maintain an element of mystery . . . ."

from second stanza "Christopher Columbus"
(W. Hart-Smith)

Chapter 20

LEARNING, ADAPTIVE AND GOAL SEEKING SYSTEMS

20.1 Introduction

An example of a system of moderate complexity has already been given in Chapter 17 with reference to a radar defence system. With such systems, the practical problems stem from our inability to predict with precision the future trajectory of the target. Nevertheless, an interesting possibility exists that given sufficient storage facilities associated with the computer we might be able to scan the data received and discover a definite pattern of behaviour. This pattern would be associated with a definite path of the projectile and our strategy of hunting the enemy aircraft would be thereafter greatly simplified. In fact from there onwards a 100% success could be assured. We note that to ensure a success the system must be programmed so as to be capable of "learning". This example serves to illustrate a system which, though variable, operates according to definite patterns.

Frequently the target will not follow a definite pattern at all, and under such conditions the data representing the position of the target will not follow any definite pattern but is more likely to represent a stochastic process. The problem is then more difficult: we have to decide, in view of the delay in transmission and the nature of the data received, on the best strategy for anticipating the future position of the target. The system must therefore be equipped with facilities suitable for processing statistical data and for extracting the information therefrom. Hence follows the relevance of the statistical theory of communication to this kind of problem.

Characteristic of systems of this type is that during the period of activity there is a flow of data (information) between the components, parts or functional units of a system. Information theory is concerned with this flow of information, and its relation to the overall performance of the system and the goals envisaged. Even with systems of moderate size there is an immense variety of ways in which various components and functional units of the system can be inter-connected. The problems associated with optimum design of systems are therefore of extreme complexity, the ultimate aim being to design a system which can reach the goals in some optimum manner.

In view of the implied complexity of the problems associated with design of systems, a systems engineer needs models to help him to comprehend the problems involved. Models can be mathematical or analogue*, but to have a useful model the problem needs to be formulated. However, a complex system problem cannot be adequately formulated until it is well understood, presumably by means of suitable models and at the same time it cannot be adequately understood until most of the problems associated with the system have been solved. This, therefore, is the dilemma of the systems engineer.

Systems theory can be extended, through interaction with other disciplines, to include human operators as parts of a complete system. From the information theory point of view, however, such extension is trivial in so far that the human operator is usually replaced by an input/output device having a mathematical expression which describes his behaviour.

The present chapter is devoted to the examination of some of the problems with systems of this nature.

20.2 Adaptive Control Systems

Essential to the functioning of any automatic system is some form of a sensor which senses the output of the system and feeds the information to some other parts of the system to make the system behave in accordance with a built-in criterion.

A feedback system as used in automatic control is a particularly simple illustration of such an arrangement (See Chapters 10 and 16). The aim here is automatically to sense and correct deviations from a chosen state. Fig. 20.1 represents a block diagram of such a system.

Such systems can be made very accurate in their steady state behaviour. However, to ensure that their transient response is well behaved can be a major problem. The transient behaviour (i.e. the response to a sudden change in circumstance followed by the efforts of the control system to cope with this) will depend on factors such as

(1) The properties of the plant

(2) The previous state of the output

(3) The nature of the disturbance

(4) The properties (i.e. parameters) of the controller

* By analogue we mean that a hydraulic machine or mechanical device can have an almost perfect model in a suitable electrical circuit. Such an electrical circuit would be called the analogue of the corresponding hydraulic or mechanical system.

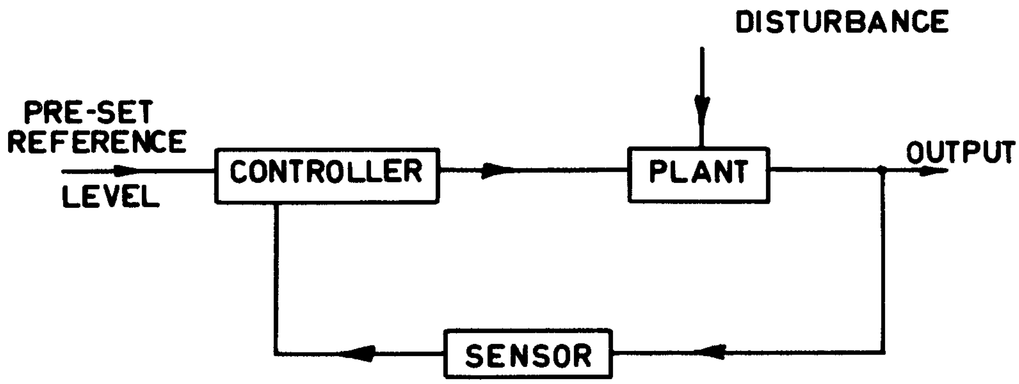

FIG. 20·1.

In view of the multiplicity of the factors involved, it is unlikely that the transient behaviour will always be optimum.

At the same time what is meant by "optimum" could be judged in a variety of ways e.g.

(1) Shortest time (to settle down to the new steady state)

(2) Least expenditure of energy (to achieve the new steady state)

(3) Avoidance of undesirable stresses in a machine

(4) Avoidance of overheating a component or device.

It is possible for a human operator to make judgments of the above sort and to decide to alter the characteristics (parameters) of the controller in order to do better in some new situation (i.e. to achieve something nearer to "optimal" control)

It would also be possible if we knew just how the human operator goes about making such judgments, to program this into a computer (either analogue or digital or a hybrid combination of both) and to arrange for the computer to adjust the parameters of the controller to achieve "better" control of the plant.

Figure 20.2 illustrates a more ambitious attempt at solving the control problem. When this system is compared with that shown in Fig. 20.1, we observe that one block, labelled computer, has been added. The connections have been arranged so that the computer inputs inform it of all the significant variables in the rest of the system. In the example shown these are

(1) the pre-set reference level

(2) the controller output

(3) the state of the plant output

(4) the disturbances to which the plant is being subjected

The output of the computer is shown by several (three, in Fig. 20.2) connections to the controller. Although not shown on the diagram it is arranged that these inputs from the computer, will vary the properties (i.e. the parameters) of the controller.

Such a system would be called an adaptive control system because it adapts its parameters and thus attempts to improve its performance. However, it is necessary for the designer to anticipate the nature of the possible deficiencies (i.e. the way in which control could fall below "optimum") and to ensure that these can be reduced by the "programme" he writes for the computer which thus decides the manner in which the computer's answers will be used to vary the controller characteristics.

In some cases the computer might only collect data from as few as two points. In other cases the computer might be fed with data collected from several dozen points.

Programmed computer control of complex systems such as the one described with reference to radar defence systems would appear to be the only practical solution to the many problems facing the designer.

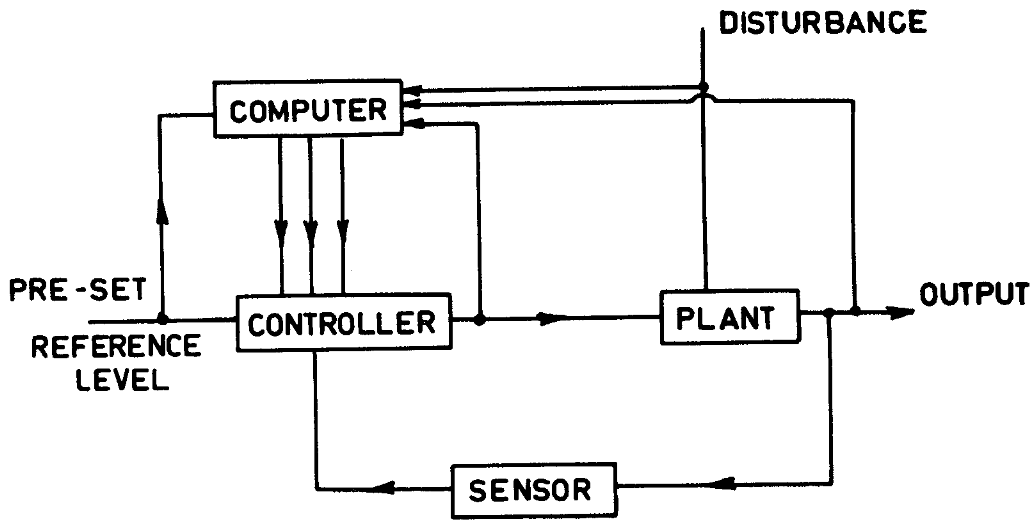

FIG. 20·2.

20.3 **Predictors and Learning Machines**

A system of the type described in the previous section needs, to function properly, to be able to anticipate the immediate future of its inputs. The devices which enable the system to make anticipatory actions came to be known under the name of "predictors". The term came into use during the Second World War when system theory was applied to the design of controllers for anti-aircraft guns. The device, a "predictive filter", enabled the guns to be directed to the anticipated position of the target. Such action was possible in consequence of calculations made on the past recorded manouvres of the target, in spite of attempts on the part of the target for random evasive action. The theory which enabled scientists to carry out such work grew out of Wiener's work. Some of the results of Wiener's work were strange and led to some misconceptions, particularly as regards the existence of devices capable of predicting future events.

A word of explanation is therefore in order. Clearly, no information-bearing characteristics or traits of any signal can possibly be predicted; this is axiomatic in view of the earlier discussions: information is equivalent to entropy which is uncertainty as to what is going to be transmitted. The part of the waveform or the signal which can be predicted cannot therefore be information in the information theoretic sense. "Prediction" as used here, refers to some stationary characteristics of the class of signals received, or the characteristics of the system as a whole. For this reason it is possible to predict the voltage value of the waveform, but on a time scale which is limited (roughly of the order of the reciprocal of the bandwidth of the system).

In the same way, it is possible to predict the position of the moving target such as an enemy aircraft, but only on a time scale which is compatible with the manoeuvreability characteristics of the target. It is not possible to predict noise voltage of a waveform at the next sampling point, nor the acceleration of the military target at the next instant of time. The rate at which a sudden acceleration can be translated into non-predictable changes in position is limited by the inertia and moments of inertia of the aircraft. While the acceleration cannot be predicted, the velocity can be for a short time, and the position for a somewhat longer time.

Whereas information cannot be predicted, it is possible to say something about the future behaviour of various integrals which are related to the information bearing variables (see experiment at the end of Chapter 4): this is a different proposition to predicting the unknown, the information.

Thus, it becomes apparent that there exists a very intimate relation between system theory and the theory of random (or stochastic) processes. Normally, the independent variable is time and the random variable can be quantities such as voltage or position of the target or the economic data pertaining to the functioning of a commercial enterprise etc. In all such cases, one of the purposes of the computer is to record the past events and to perform various calculations on the recorded data so as

to deduce various statistical characteristics of the input signals. From the data so obtained, the computer subsequently evaluates the various statistical averages such as the mean position of the target, the expected direction of the movement of the target etc. Thus, we see, it is these various statistical averages which can be predicted, and the control system is acting in accordance with such statistical data so as to optimise the performance of the system as a whole. The statistical data so obtained virtually characterises the environment in which the system is operating. It is this statistical data in various processed forms which is used subsequently for decision making purposes.

In addition we will need some arrangements* to allow communication between machine and man and between machines.

20.4 Variable Systems and Loss of Information

There is an aspect of machine-environment interaction that needs a little elaboration concerning the so-called variable or time-varying systems (c.f. example of radar tracking system). These systems can be modelled in several different ways, but the capacity of a variable system can be expressed quite simply using a slightly extended concept of entropy (see Chapter 4). In particular, if the variation in the properties of the system can be described by a definite mathematical relation, however complicated, then in absence of noise the capacity of the system will not be affected at all, unless the time variable operator representing the system is a many-to-one operator.

This is because under such conditions the entropy is unaffected. In engineering terms, the particular mathematical relation could be stored in the machine and used in the encoding process without affecting the information carrying capacity of the system. †Yet, the form of the presentation of the signal would be altered and could, through changes in the semantic content, confuse the observer.

Again, in the last example, distinction must be drawn between the different aspects of a signal: (i) its entropy, and (ii) its form of semantic content. Whereas the second is of no concern in communication theory, the first one is.

If the system were variable in a stochastic sense then clearly there would be a corresponding entropy associated with it, H_s. Under such conditions, the capacity of the system would be reduced to

$$H = H_o - H_s \qquad (20.1)$$

where H_o denotes the capacity of the system in the absence of the stochastic variations within it.

A further point worth observing as far as variable systems are concerned, is that such a system can be regarded as a constant system provided with an additional input port. Clearly, in such equivalent presentation, if the additional input ports contain signals which can be described by definite mathematical relations, then the system capacity will be unaffected; but, if the inputs to the abovementioned ports are noise waveforms then the capacity will be reduced in accordance with Equation 20.1.

At this juncture it is helpful to note that from the point of view of communication theory, a variable system, in the presence of statistically constant source (stationary environment), is equivalent to a constant system with statistically variable input (non-stationary), provided that the balance of entropies in the two cases is the same. Thus, a change in the form of the input signal is unimportant, but a change in the *a priori* and *a posteriori* probabilities is.

A variable system, or a constant system operating in a variable environment, needs for its proper functioning an adaptive mechanism: A "predictor" is its essential constituent. An example being radar tracking system. The system as mentioned previously cannot predict the information, (See Section 4.4) it can only collect data and adapt itself to conform to an optimum criterion based on statistical averages and the information for this must come from the entropy of the received signals. The capacity of the system is therefore reduced and, in a sufficiently quickly varying environment the whole of the system's capacity might be occupied by the entropy needed to up-date the statistical averages: This is the breaking point of the system. For example, with radar tracking system, the breaking point is reached when the standard deviation** of the target exceeds the radius of influence of the projectile. This could be on account of delays in transmission (capacity too small) or due to increased manoeuvrability of the target (rate of transmission of independent data too high).

* In the present-day jargon of computer engineers these arrangements are sometimes known as peripheral equipment or interfaces. These terms came into usage in that order and indicate a change in emphasis. The use of the word peripheral indicates a pre-occupation with the computer as the central feature of a system. The use of the word interface connotes more emphasis on the system as a whole.

† But, if such mathematical relation is not known or its implementation would exceed the storage capacity of the system then this route is not open to exploitation. Such mathematical "information" is no asset.

** The standard deviation or variance of a set of numbers $x_1, x_2, \text{---------} x_n$ from their mean value $\bar{x}$ is defined as the average of all the terms $(x_i - \bar{x})^2$ where i takes on all the values from 1 to n successively. Notice that the positive and negative errors do not cancel out because the square of the error is taken each time. A set of widely scattered readings will have a larger variance than a closely clustered set.

Frequently, it happens with adaptive systems that the reason for reaching a breaking point is not the lack of information capacity per se but rather that the form (or semantics) of the signal is a function of time and it is the adaptive system (dealing with data) that interprets the signal as highly time dependent. In such cases a self-organising system – which can change its characteristics as the need arises – has a clear advantage.

An effective implementation of such devices (excepting some very simple cases) hinges on the discovery of efficient means of dealing with semantic characteristics.

20.5 **Patterns and semantics**

All learning and adaptive machines have facilities which enable the systems to perform pattern recognition, and there are a number of facets of varying degree of complexity.

In the first place, if we know the constraints then we can calculate the total number of distinct patterns. However, we might encounter several difficulties: the constraints may not be completely known or be too complex for adequate description, or the total number of possible distinct patterns may be prohibitively large.

For example, the class of patterns may be specified as that of sinusoidal waveforms that is

$$\Psi_n = A_n \sin (\omega_n t + \phi_n) \tag{20.2}$$

$$\text{Where } \Psi \triangleq (A_n, \omega_n, \phi_n) \tag{20.3}$$

The problem of pattern recognition reduces to the determination of three numbers.

One of the functions performed by a machine may be pattern identification, where we ascribe to every pattern of the set a symbol (the name of the pattern) on a one-to-one correspondence. This is a problem of encoding and clearly possesses a solution, but it may be complicated in that it might be difficult to implement in practice, unless a systematic method is conceived.

To illustrate the point, in the example given the set of three numbers (A, ω, ϕ) identifies a pattern out of a class of sinusoidal functions, on a one-to-one basis.

Pattern classification is another function which machines are frequently called upon to perform. Pattern classification is another example of encoding, whereby we ascribe one symbol to a group of patterns. As such it is an example of a many-to-one translation. The information necessary to accomplish the correct pattern classification into the various groups, is either built into the machine (for operation in a fixed environment) or is "learned" by the machine with the help of a "teacher". In this way the machine can be used or can be made to adapt to different environments.

Returning to the example given above, the machine might be asked to classify all sinusoids of frequency greater than ω_0 as X and send them to port A and all other waveforms as Y and send them to port B.

One could regard such a machine as exhibiting an intelligent behaviour, particularly so if by manipulating suitable knobs a "teacher" could "teach" the machine to adjust the selection point, ω_0, to a desired value. On reflection however, one sees that such a machine is just as embodiment of a single two-output-port (X & Y) filter with none of the fascinations of a learning machine, and man (teacher)-machine interaction. One needs, therefore, to be careful when using terms with emotional connotations, such as intelligent machines or thinking machines.

Single detection (with communication systems) is another example of a function following in the class of pattern classification schemes in that the process concerns separation of mutilated patterns into different classes according to a prearranged plan.

The purpose of a machine performing pattern recognition or classification functions is to reach a goal according to a prearranged plan. This plan is essential for the successful operation of the system. As such, it is an attempt at translating patterns present in the minds of planners (e.g. engineers) into a machine language, and the process involves measurement as an essential step in the execution of the various functions by the machine. It follows, therefore, that the machine can classify only such items which are amenable to quantitative description. Frequently, it is this step - i.e. conversion into a quantitative or digital form – that is a source of difficulty. Thus, emotions and feelings which are the essence of human communication cannot be communicated effectively by a machine directly, simply because we know of no suitable scale of measures.

Much of human communication is semantic in content and whereas communication theory deals with entropy of patterns, humans are frequently more concerned with terms like simple, complex, beautiful etc.

What is simple for a human being is frequently difficult to implement by a machine and vice versa.

Thus it appears that whereas humans are particularly adapt at classification according to form (or semantics), machines are more suitable for dealing with quantitative . . . (Cont. next page)

data. In this sense machines complement humans, but a great deal needs yet to be done to solve the interface problem; that is, effective methods of translating semantics into a quantitative language.

Frequently, pattern classification can be accomplished by machines as well as by humans but the strategy employed by humans appears to differ fundamentally from that used by machines. Humans, in particular, have a remarkable ability for pattern recognition in that they can select at a glance out of a maze of billions of possible alternatives, those which are "worthwhile" thereby reducing the search procedure considerably. (But note that "worthwhile" implies emotional involvement. None of the machines invented to date have such capabilities).

Pattern recognition is a province of information theory, irrespective of the effectiveness of the identification procedure, or strategy. This is the same class of problems as detection of signal (given class) in noise. However the problem of establishing whether there is a signal (class not given) in a noisy background or whether there is a pattern (class not specified) on a radar screen, is a much more difficult problem and often outside the province of communication theory.

Thus it appears, at least in principle, that given a sufficiently large and complex system we can classify patterns, *and* detect them in noisy background, and we can design a system to adapt to a changing environment even though the machine might do it in a rather clumsy way. But the problem of devising a system which would search for a pattern, as yet unspecified and, having discovered it, would take appropriate action would appear to be beyond our present capabilities. The fundamental question to which, in the course of discovery, a scientist frequently seeks an answer is "does this series of measurements mean something", or alternatively "is there a pattern or a law in this otherwise random set of data?" While history shows clearly that the bulk of great discoveries in science came in response to such questions, yet no machine could answer such questions (being semantic they are not amenable to quantitative scrutiny). The questions are not "scientific", they imply strong emotional involvement.

It is this constant search for elusive patterns which makes science and engineering so fascinating.

20.6 Future Uses of Learning Systems.

We are only just starting to learn how to set about designing machines that can learn. This ability when it is achieved will create a new technological revolution.

Over the last couple of centuries we have seen the design and widespread construction and use of machine which have, in a sense, extended the capability of human and animal muscle power.

In the present century we will undoubtedly see the design and widespread use of machines which will extend the power of the individual human mind.

It will be a very grave political problem as to whether the ownership and control of such machines should be in government or private hands or in both.

It will be a very grave sociological problem to decide under what terms an individual or a corporate group (such as a large firm) should have access to or should monopolize such learning machines.

In order to analyse and understand such machines it will be necessary to be acquainted with a much wider range of mathematical processes than engineers have been accustomed to use the past. Some of these are included in the following list and are in addition to all those techniques which we have previously mentioned in connection with other sorts of systems.

Group theory.

Variational calculus.

Symbolic logic (including the use of Boolean algebra for processes other than simple binary logic).

Functions of discrete variables (i.e. compared to continuous variables.)

Field theory of discrete sets (i.e. compared to field theory of a continuum)

These and other areas in mathematics will be needed in order that we can deal with both digital and analogue processing of data in machines which provide recognition, memory, association, re-enforcement and inference.

Systems theory in its broadest sense is the tool by which engineers, social and political scientists, military strategists and tacticians, commercial, industrial and agricultural managers, local, national and supra-national planners and executives may

all play their role in the years and societies yet to come. It may be a tool which will influence humanity for good or ill as drastically as the industrial revolution or the discovery of atomic energy.

Looking back on the industrial revolution which provided a great increase of energy per capita, one may see that the actual utilization of this energy remained however, by and large, under individual human control. One may ask, was it coincidental or was it necessary that mass production arose and prospered at the same time by subdivision of manual tasks into almost trivial repetitive processes.

The question must arise, will the present-day and future applications of systems engineering induce a similar subdivision of human perceptual and mental tasks into sets of strictly-controlled trivia. One hopes and believes that this will not be so. Indeed the problem of ensuring that the new engineering, which is ushering in the twenty-first century, should truly enrich rather than impoverish human experience is indeed itself a problem in systems engineering of the greatest magnitude.

20.7 The global system: a plan for to-morrow

And what of the future? Is there no limit to the complexity and the sophistication of systems?

The industrial revolution has made impacts in its time both desirable as well as undesirable. Systems engineering is beginning to make its impact on the present century. Like the industrial revolution, systems engineering, while benefiting mankind, could have far reaching highly undesirable effects: we can already foresee some of the problems ahead, and should be concerned with them now. Let us examine a few.

Large systems are expensive. Any particular system may represent to a nation a substantial investment, and the whole economy of the country could be very profoundly affected by the malfunctioning or the failure of the system. Moreover, a significant downward trend in the economic situation of one country might have profound effects on others.

The environment, the world at large, is not an inertial frame of reference in which different systems could be said to operate substantially independently. But rather, like the building blocks of a system which form the whole (see Chapters 9 and 10), so the various sytems are *not independent* but are *coupled* to each other through the medium of the environment. With large systems this coupling can profoundly modify the functioning of the system in a manner which may not have been foreseen by its designers. It would even be possible for the system to become unstable.

Furthermore, the coupling forces could be non-physical and be related to human aspects. As such they could not be measured and taken into account in an unambiguous calculable manner.

To make the point clearer, it needs to be observed that systems as we know them today are sophisticated goal-seeking machines which tend to optimise their performance in accordance with a criterion relating to a self-centered policy in a hostile environment. It is this kind of policy which must be criticised if a number of systems are compelled to work in a common environment.

It is not difficult to show that several systems of the type described and operating in a common resilient* environment lead to an operation which is inherently unstable and therefore none of the component systems can optimize their performance. Moreover, the conclusion still stands, even if the coupling forces are small.

The corollary is inescapable: the self-centered policy leads to an operation which is bad as far as all component systems are concerned, and no one profits by it. Perhaps a profound lesson can be learned from this. Are we thus doomed to disaster?

The answer, fortunately is "No", provided that the self-centered folly is abandoned and that a global optimisation policy is substituted. This works as follows. The various systems together with the interaction forces are considered on a global basis, as a global system, so as to optimize the performance of the global system without it being unduly concerned about the local value of the pay-off of a particular sub-system. The benefits accruing from the global system would be re-distributed in an agreed fair way. There would be no hostile environment, and the system would be inherently stable.

Undoubtedly global systems, whose operational criterion is that of genuine concern for all, would benefit mankind enormously: the alternative is global disaster.

Which is it going to be? The latter, the blunder of ignorance being the reward for blind self-centered policy, pre-occupied with short-term gains, or the former global system based on a policy of genuine concern for each and every one of us, a triumph and a living monument to education.

"The earth is but one country and mankind its citizens."

Baha'ullah.

* by a resilient environment we mean one which substantially conserves (rather than dissipates) the coupling forces (c.f. the terms reactance and resistance used in electric circuit theory).